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The *b* of the van der Waals Equation

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IN the handbooks and advanced textbooks there are found two elementary methods of deriving the correction for the finite volume of the molecules in the equation of state of a gas. The connection between the two is not apparent. Those who have taught the subject know that the topic becomes less than crystal clear in the minds of students who venture to study more than one derivation. This is not their fault, I believe. One of the common treatments seems to be fallacious, and the other in need of more critical consideration than is usually given.

The following two quotations are illustrative of the first type of derivation. I do not know who originated it.

> "Firstly, we must subtract from the macroscopic volume v a quantity b which allows for the space occupied by the molecules. This quantity b is four times the volume actually occupied by the molecules, which are assumed to be spherical. The reason for this is that in a collision the centers of two molecules approach no nearer than a diameter, so that for impact we may imagine half the molecules to be mere points, while the remainder occupy eight times their actual volume. Hence, for the average, we use four times the true volume of the molecules."

> "The radius of the molecule may be determined from the constant b in the van der Waals equation. This constant represents the portion of the volume occupied by the gas which has to be excluded from the phase space on account of the size of the molecule. We shall find the relation between it and the radius a of the molecule on the assumption that simultaneous collisions of three or four molecules are so rare that we make no appreciable error in proceeding as if all collisions occur between two molecules alone. If two molecules collide, their centers get no nearer than the distance $2a$. So a space $(4/3)\pi(2a)^3=(32/3)\pi a^3$ must be omitted from the volume v on account of a pair of colliding molecules, or $(16/3)\pi a^3$ for each molecule. Therefore, if N is the number of molecules in a volume v, $b=(16/3)\pi a^3N$, which is four times the space actually filled by the N molecules."

These arguments are supposed to be based on elementary principles. They are given in books intended for the use of students without much background from previous study of such matters. They seem to imply that it is obvious that a swelling of infinitesimal molecules to finite size would increase the pressure just as it would be increased by introducing a solid body into the space open to the molecules. This is far from obvious. Pressure arises from the impact of the molecules at the walls and is proportional to the product of the mean-square velocity and the concentration of the molecules in the neighborhood of the wall. Suppose N molecules of negligible diameter to be in a space of volume V. Let them swell up to a diameter σ. The mean number per cubic centimeter will still be N/V as before.[1] The mean velocity is still the same, granting the same temperature. How is any increase of pressure to arise? This line of argument seems to be leading to the erroneous conclusion that there would be no increase of pressure. It is necessary to consider the effect of collisions in increasing the rate of propagation of momentum with resulting increase of fre-

[1] This leaves out of account the entirely negligible decrease of the effective volume resulting from the fact that the center of a molecule can approach no closer than $\sigma/2$ to the wall.

quency of impact at the walls, as was done by Clausius.[2] Such considerations are relatively sophisticated and need to be given explicitly if it is intended that they constitute part of the argument. The treatments quoted seem to be so incomplete that the conclusion as to the quantitative value of the increase of pressure is a *non sequitur*, and the argument therefore fallacious. To make it a proof it needs first, a setting forth of the reasons why an increase of pressure is to be expected, and second, a deduction of the numerical value of the increase on the basis of those considerations.

The other proof frequently encountered is that of Boltzmann.[3] It is given in fairly complete form by Jeans[4] and need not be repeated here. The important point which is brought out clearly is that the concentration of molecules, when they are of finite volume, is actually higher near the walls than elsewhere in the gas. Jeans points out that the argument of Boltzmann is subject to some criticism as to the way in which the theory of probability is applied. He does not consider the matter further, as he is about to derive the same result by more general methods. The same argument made with only two molecules brings out some of the essential points.

Assume a vessel of volume V' into which two molecules, each of diameter σ, are to be put. Since the center of neither can come closer to the wall than $\sigma/2$, the effective volume open to centers of the molecules is less than V'. Call it V. Put in the first molecule. If we assume that one position is as likely as another the chance that the center will lie in any element of volume of size dv will be dv/V. Let the second molecule be put in and consider the probability that it will lie in some element of volume dv. This probability is equal to the product of the independent probabilities that dv is not within the sphere of exclusion of the first molecule, and that if dv is not in the sphere of exclusion, the center of the second molecule will lie within that element dv. The probability that any element dv out in the middle of the vessel will be shielded by molecule No. 1 is evidently $(4/3)\pi\sigma^3/V$. The probability that it will not be so shielded is $1-(4/3)\pi\sigma^3/V$. If the second molecule is likely to be anywhere in the space available, the chance that its center will lie in dv is $dv/[V-(4/3)\pi\sigma^3]$. The probability that the center of molecule No. 2 will lie within dv is the product of these two last probabilities, which gives dv/V. The result is the same as if the size had been neglected. If, now, dv be taken as close to the wall as possible ($\sigma/2$ away from it) the probability that dv will be shielded by molecule No. 1 drops to $\frac{2}{3}\pi\sigma^3/V$, one-half of its former value. (No shielding molecules can be on the back side of dv.) The final probability is accordingly

$$\left(\frac{dv}{V-(4/3)\pi\sigma^3}\right)(1-\tfrac{2}{3}\pi\sigma^3/V).$$

To the first order of small quantities this equals[5] $dv/(V-\frac{2}{3}\pi\sigma^3)$.

Summarizing then, the probability that molecule No. 2 will be found in an element dv further than σ from the wall is dv/V, the probability that it will be in dv close to the wall has the larger value, $dv/(V-\frac{2}{3}\pi\sigma^3)$. For intermediate positions it will have an intermediate value. If the probability now be summed for all elements to get the probability that molecule No. 2 will be somewhere in the vessel it comes out to be greater than unity. This unhappy result is merely the just reward for the careless way in which the foregoing argument was made (and usually is made). As soon as a second molecule is put in, it is no longer legitimate to assume that all positions are equally likely for molecule No. 1. If there is a greater chance of a molecule (either one) being close to the wall the probability that dv out in the gas will be shielded must be slightly less than $(4/3)\pi\sigma^3/V$, and the probability that dv close to the wall will be shielded is slightly greater than $\frac{2}{3}\pi\sigma^3/V$. It is also clear that such changes made in the interests of logical consistency would produce only changes of the second order in the final result so that the result of the Boltzmann argument is evidently correct to the first order of small

[2] See, for example, the article by Herzfeld in the Mueller-Pouillet *Lehrbuch der Physik*, Vol. III (Zweite Haelfte) p. 50.

[3] Boltzmann, *Gastheorie*, Vol. II, p. 7.

[4] Jeans, *Dynamical Theory of Gases*, ed. 3, p. 126.

[5] In the Boltzmann proof with N molecules the corresponding result is

$$\frac{dv}{V-(N-1)(4/3)\pi\sigma^3}\left\{1-(N-1)\frac{\frac{2}{3}\pi\sigma^3}{V}\right\}\cong\frac{dv}{V-N\cdot\frac{2}{3}\pi\sigma^3}.$$

The number of molecules in dv is taken to be this probability multiplied into the total number.

quantities. This is all that it purports to be.

The same result comes as a first approximation from the virial theory, and from statistical theory. These latter methods are certainly the more elegant ones. There are advantages, however, in a theory which is more pictorial, especially when the picture is not too fictitious.

I am greatly indebted to Dr. Karl Herzfeld, whose thorough criticism of the first draft of this paper clarified my thinking in the matter considerably and led me to make several changes. He must not be held responsible, however, for any errors that there may be in the paper. Those I must cling to as my own.

The Untold Story of the Telephone

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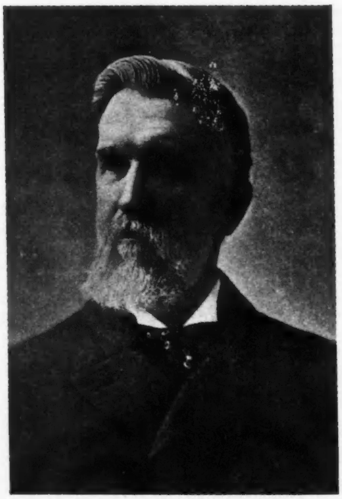
ON December 13, 1934 the newspapers noted the death of Thomas A. Watson, the man who, as the assistant of Alexander Graham Bell, was the first American to understand a sentence spoken over the telephone. With his demise went the last survivor of a small group of men who had a first-hand, uncensored knowledge of the circumstances surrounding the birth of the telephone. Much of the story of that great discovery they have given to the world. Other portions of it which they have left untold are perhaps of equal interest. Some of these untold portions are of record, though the records are widely scattered and there has been a surprising lack of competent endeavor to bring them together and integrate them with the more publicized parts of the story. Some of the less well-known but more important portions are to be found in court and patent office records, in technical publications, in the proceedings of an investigating committee of the United States House of Representatives and in memoirs and personal correspondence of some of the principals. While it is unlikely that even the sum total of all these sources embodies the entire story of the telephone, it does bring out some points that place the publicized portions in a very different light than that in which they are commonly viewed.



Elisha Gray, as he appeared when he organized and presided at the first International Electrical Congress in Chicago in 1893. [From *The Electrical Journal*, Sept. 1, 1895.]

The prevailing tendency to attribute great inventions exclusively to single individuals may be natural. It is certainly fostered by the way in which our patent laws are formulated and administered, but it creates a major problem for those who are interested in historical and scientific accuracy. In the sciences allowance is regularly made for the possibility of independent discoveries, whether simultaneous or not, and there are many important chapters of science in which that element is prominent. The impossibility of sharing credit within the framework of our patent law creates a prolific source of distortions of historical fact. The identification of Alexander Graham Bell with the birth of the telephone, to the exclusion of all others, is one example of this tendency. As in many other instances, the Bell tradition greatly oversimplifies the actual circumstances. It had its birth in a skilful combination of truth and silence out of which has grown the mighty myth that Bell was primarily responsible for the invention of the telephone.

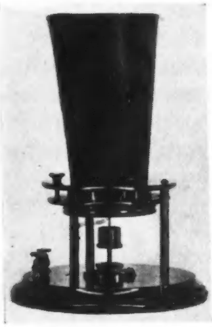


FIG. 1. The transmitter which Bell used on March 10, 1876.

More than sixty years have passed since the momentous day in March, 1876, when Bell spoke the first sentence ever transmitted in this country by electricity. That exclamation, "Mr. Watson, come here, I want you!" is justly celebrated as the beginning of a new era in communication. The episode is fully authenticated; nor is there any good reason to suppose that it was antedated in this country by any other successful attempt of the same kind. The story has been told many times by the participants themselves. The most recent account is to be found in Watson's autobiography.[1] He describes the occasion as follows:

> It was during one of Bell's experiments on this kind of a telephone that the first sentence was transmitted and understood. I had made for Bell a new transmitter [Fig. 1] in which a wire, attached to a diaphragm, touched acidulated water contained in a metal cup, both included in a circuit through the battery and the receiving telephone. The depth of the wire in the acid and consequently the resistance of the circuit was varied as the voice made the diaphragm vibrate, which made the galvanic current undulate in speech form.
>
> I carried the transmitter when finished to Exeter Place on the evening of March 10, 1876, intending to spend the night with Bell testing it. Neither of us had the least idea that we were about to try the best transmitter that had yet been devised. We filled the cup with diluted sulphuric acid, and connected it to the wire running between the two rooms. When all was ready I went into Bell's bedroom and stood by the bureau with my ear at the receiving telephone. Almost at once I was astonished to hear Bell's voice coming from it distinctly saying, "Mr. Watson, come here; I want you!" He had no receiving telephone at his end of the wire so I couldn't answer him, but as the tone of his voice indicated he needed help, I rushed down the hall into his room and found he had upset the acid of a battery over his clothes. He forgot the incident in his joy over the success of the new transmitter when I told him how plainly I had heard his words; and his joy was increased when he went to the other end of the wire and heard how distinctly my voice came through.

There has never been any occasion for questioning the story as thus told by Watson. He and Bell have a clear priority to the claim of having conducted the first American telephone conversation. It is true that this story was not publicly told for several years after the event, but this was due, not to any suppression of it by opposing interests, but through the apparent intent of Bell himself. The most interesting feature of the episode is, however, the fact that Watson's account constitutes only a part of the story. There are some aspects of the story as a whole, now almost forgotten, which place it in an entirely different light than that in which it is customarily presented. For the entire story to have come to light would have been a catastrophe from Bell's point of view. Hence it may be that the delay in telling it arose from the necessity for waiting until the selection of certain portions of it and the discarding of others could be made with less danger of challenge.

Four of the most interesting elements in the setting of Watson's story which do not appear in that or any other of the official versions are the following:

> *First:* that the transmitter into which Bell spoke on March 10, 1876 was a very different kind of instrument from that described and illustrated in his earlier patent.
>
> *Second:* that the transmitter which he had constructed for the occasion had previously been described by a man named Elisha Gray in a confidential document about the

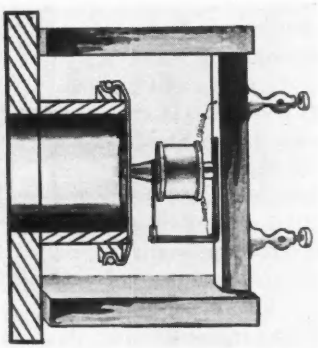


FIG. 2. Cross section of Bell's transmitter of 1875. Later used also as a receiver.

[1] *Exploring Life* (D. Appleton-Century, 1926), pp. 77–81.

contents of which Bell subsequently acknowledged having received information.

Third: that it was not until four years later that Bell made any claim to the type of transmitter into which he spoke on that historic occasion.

Fourth: that Elisha Gray had made and publicly used several types of telephone receiver many months before Bell constructed his first one.

These theses may now be developed somewhat more fully. The first three deal with the microphone type of transmitter, which is now universally used, and the fourth with the metal-diaphragm type of receiver, equally common. Both are commonplaces today, but neither was a commonplace in 1876. The telephone controversy centers primarily in the transmitter.

The first point is, then, that *the transmitter into which Bell spoke on March 10, 1876, was a very different kind of instrument from that described and illustrated in his earlier patent.* Bell's exclamation to Watson had the great historical significance of being the first words of any American experimenter to be spoken into a microphone type of transmitter. Hence it was the first telephone conversation, if a one-way exclamation can be termed such, in which the prototype of the modern transmitter played a part.

Now Bell had already invented a telephone transmitter, and several months prior to his famous exploit of March, 1876, had made attempts to operate it.[2] A copy of the original, made under the direction of Bell himself, is in existence today. Fig. 2 is a drawing of the cross section, showing the instrument in position for use. During Bell's 1875 experiments, sounds had come through and been heard at the distant receiver, but no words had ever been understood. Moreover, and this is the significant point, the transmitter used in these earlier trials, and the transmitter described and illustrated in his patent of early 1876, was *not* a microphone type of transmitter, but was simply a receiver, of electromagnetic type, worked backward, now a familiar device in the physics laboratory. The discovery that a receiver can be so used originated with Bell. He is clearly the inventor of the

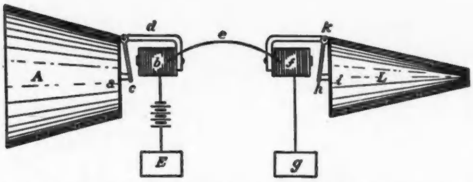


FIG. 3. Transmitter (left) and receiver (right) illustrated in Bell's first telephone patent, granted March 7, 1876.

electromagnetic type of transmitter. The point to be borne in mind is that his first successful telephone conversation was conducted, not with that type of transmitter, but with the so-called liquid transmitter, described by Watson, which can be recognized as a microphone type, the precursor, however rudimentary, of the modern transmitter.

That the transmitter which was described and illustrated in Bell's patent,[3] the application for which was filed on February 14 preceding his famous exploit of March 10 in the same year, was not a microphone type, but an electromagnetic type (Fig. 3), the only drawing of a transmitter in Bell's patent constitutes an indication. The descriptive material accompanying it treats quite explicitly its construction and operation. It is also true, however, that Bell refers briefly in his patent to the possibility of a liquid transmitter, a reference confined to eight lines out of a total of about 190 printed lines. The legal status of this passing reference has never been clarified. To the lay mind, and especially regarded as a scientific accomplishment, it will be natural to feel that mere reference to a possible invention does not itself constitute an invention. That is all that can be said for Bell's claim to the liquid transmitter. Even from the legal side, moreover, it is well to note that Bell's later victories in litigation did not center in any claim that he may have made to the liquid transmitter. The main issue in all the suits in which this patent was involved was a different paragraph ("claim") not connected in any way with the liquid transmitter. So far from emphasizing the liquid transmitter, Bell explicitly stated his preference for the receiver type of transmitter in his application. Indeed, he thought so little of the liquid transmitter that he made no reference to it in another version of his patent

[2] For Bell's own account of this, see pp. 67–70 of *The Bell Telephone,* printed in 1908 by the American Bell Telephone Co. This is a reprint of Bell's deposition in connection with a suit once brought by the United States to annul Bell's patent. For Watson's account, see reference 1, pp. 66–71.

[3] No. 174,465 of Mar. 7, 1876, entitled *Improvement in Telegraphy.*

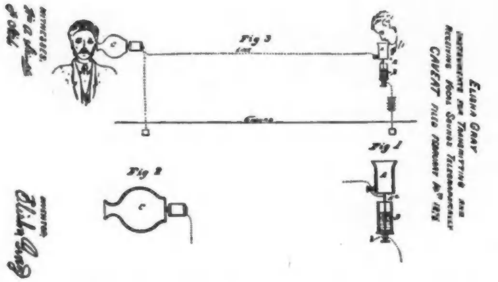


FIG. 4. Transmitter (right) and receiver (left) of Gray's caveat, filed the same day as Bell's application for a patent.

application, which he sent to England *after* he had written the American version, in the hope of securing foreign telephone patent rights as well.[4] So the transmitter into which Bell spoke on March 10, 1876 was not only not the instrument described and illustrated in his earlier patent, but it was not even the instrument of Bell's preference up to and indeed long after, his experiment of March 10, 1876.

The second thesis is that *the transmitter used by Bell on this famous occasion had previously been described by Elisha Gray, a free-lance inventor in the telegraphic field, in a confidential document about which Bell subsequently acknowledged having received information.* On the same day that Bell filed his application in the Patent Office in Washington, February 14, 1876, his principal competitor in the inventive field, Elisha Gray, filed another document, describing a telephone. Though Gray's description was fully as explicit and complete as Bell's, it was not technically an application for a patent. It was what was known as a *caveat,* a type of document which no longer exists in patent law, which was in essence a formal notice of intention to perfect and file a patent application. It gave the claimant documentary evidence of priority of an idea in event of subsequent disputes. There was no reason why the papers of Bell and Gray should not have been of the same type, both of them applications for a patent, or both caveats; the two men were on the same footing, neither having reduced to practice the idea which he was describing. But their legal advisers apparently had each his different mode of approach. As it proved, Gray was badly advised in submitting his idea in the form of a caveat. The legal technicalities involved in the two documents acted, in later court proceedings, to give Bell an enormous advantage which had no counterpart in any actual superiority or priority of his device. If both men had submitted the same type of document, whether caveat or application for a patent, the case could have been fought out on its merits instead of on the basis of a legal fiction. If Bell had submitted a caveat and Gray a patent application, thus reversing their positions before the law, we should, in all probability, have today the Gray Telephone Company in place of the Bell Telephone Company.

In addition to the difference in form, there was a difference in content of the two documents which was of even greater significance. Whereas Bell had made only parenthetical reference to the possibility of the liquid transmitter, devoting about one twenty-fifth of his application to it, Gray evidently considered the liquid transmitter to be of greatest importance. He described it in utmost detail, and illustrated it fully (Fig. 4). Half of his entire document was devoted to it. Gray's description covers the principle and many of the structural details of the transmitter made by Watson for Bell less than a month later. Fig. 5 is a photograph of a pair of the surviving models of Gray's instruments.

In pursuance of his official routine, the Patent Examiner notified Bell that his application came into conflict with a caveat which had been filed the same day. Since all such documents were confidential and the patent ethics of that time required that they remain so, the Examiner did not, of course, include in his notice to Bell any statement of the points at which the two claims interfered. Bell was acutely aware of the importance of his claims to invention of the telephone, doubtless feeling, and with the best of reason, as subsequent events have shown, that a fortune was at stake. Consequently he was deeply concerned at any possibility that this

[4] The English version was dispatched through one George Brown in the last week of January, 1876. Bell was always very vague about the date of his conception of the liquid transmitter, but his most direct statement on the point is to be found in *The Bell Telephone,* p. 87, where he sets it at "almost the last minute" before sending his patent application to Washington. This would probably have been about January 10–12, 1876. See also Catherine MacKenzie, *Alexander Graham Bell* (Houghton Mifflin, 1928), pp. 109–110.

fortune might elude his grasp. Apparently at the advice of his lawyer, he called on the Patent Examiner and asked to be informed about the nature of the interfering claims. Bell may have been entirely innocent in this move. Though he had had previous patent experience, he was not well versed in either the law or ethics of patent procedure, and may not have realized that he was asking for a breach of official confidence which would give him an unfair advantage over his competitor. He should, of course, have been instructed on this point by his lawyer. In any case, the only honorable course open to the Patent Examiner was to refuse to divulge the requested information. For some reason, however, he did not refuse. Pointing to the paragraph in Bell's application which mentioned the liquid transmitter, he told him that there was an interference at that point. What led him to do this we shall probably never know. The fact that he did it has been established by Bell's own repeated admissions in court.[5] The Examiner never appeared in court to deny it, nor did he ever make any denial outside of court. The fact that Bell received this information is completely authenticated.

Statements have been made and evidence offered to the effect that the Examiner went much further than this, allowing Bell an extended examination of Gray's caveat.[6] The allegation seems not to have been adequately proven, though it has certainly never been conclusively disproven. In any case, it is relatively unimportant. If the Examiner gave no more information than that which Bell admits having received from him, he violated the responsibility of his office. In doing so, he gave Bell just the information that would spur him to the step that he next took, construction and use of the device that his competitor had described, but never made. There is no suggestion anywhere that up to this time Bell had ever made, or contemplated making a liquid transmitter. Yet he immediately recalled his assistant, Watson, whom he had previously laid off, and through feverish activity on the part of the two, succeeded in making and successfully using a liquid transmitter within less than two weeks of the time that he received information about it through the Patent Examiner. Thus not only had Bell's transmitter, made by Watson and first used on March 10, 1876, been previously described and the description come to Bell's knowledge, but the subsequent occurrences indicate that he made good use of the information, without which his first successful use of a telephone would at least have been considerably delayed.

The third thesis is that *it was not until four years later that Bell made any claim to the type of transmitter into which he spoke on that historic occasion.* The first successful transmission of speech in America by means of electricity was an historic event. Bell might have been pardoned, indeed, he would have been fully justified, in publishing it far and wide. Today it is universally considered the most dramatic episode in the history of the telephone. It is somewhat of a shock, therefore, to discover both that Bell did not at that time make any announcement of his accomplishment and that for several years he apparently made every attempt to conceal it. Some of the principal occasions when, having an opportunity to tell of that now famous episode, he nevertheless refrained from doing so, will be recounted presently. In the meantime it may be

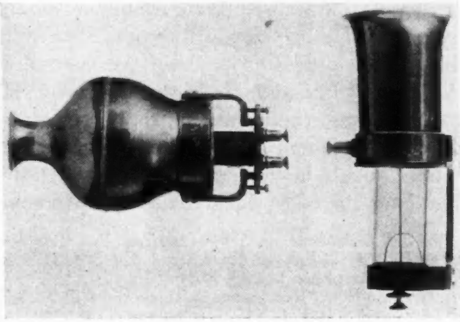


Fig. 5. Surviving models of Gray's receiver and transmitter.

[5] For example, see *Dowd Case Record* (Circuit Court of U. S., Dist. of Mass., 1880), I, 529–531; *The Bell Telephone*, p. 434.

[6] A long affidavit to this effect by Zenas F. Wilber, examiner in charge of electrical applications in the Patent Office during the interval in question, was published in *The Washington Post* of May 22, 1886. The original autograph copy of this affidavit is in the author's possession. This affidavit named five witnesses to the alleged conference and contained statements which, if untrue, could have formed the bases of several suits for libel. No such suit was ever instituted. Bell published a denial of Wilber's allegations in the May 25 issue of *The Washington Post*, and there the matter was allowed to rest.

observed that if he was to conceal the incident and yet claim the distinction of having been the first to talk over a telephone of his own make, he must reproduce the incident, using an instrument which he felt he could claim. This he prepared immediately to do. He set about the task of so modifying the receiver type of transmitter described in his patent that it would transmit speech, and within three or four weeks he was successful, in a second episode which deserves to rank with that of the preceding tenth of March.[7]

Having so altered the telephone of his patent that it could be made to speak, Bell was then in a position to make public announcement of his accomplishment of the electric transmission of speech and to demonstrate it without using a liquid transmitter. This he did in two important public lectures in the late spring of the same year, one before the American Academy of Arts and Sciences on May 10 and one before the Society of Arts at a meeting held at the Massachusetts Institute of Technology on May 25.[8] Bell may have had his liquid transmitter with him on these occasions, or even shown it to his audiences, though there is nothing in the printed record of his lectures to indicate that he did.[9] Bell himself, when testifying a few years later, could not recall whether he had the liquid transmitter with him on those occasions.[10] In any case, he did not demonstrate it in operation, though he did demonstrate his receiver-type instruments.

Bell must have felt very insecure about any claim which he could make to the liquid transmitter, for he continued for a long time this policy of secretiveness about his now famous experiment of March 10, 1876. An unprecedented opportunity for bringing his accomplishment to public notice presented itself in the Centennial Exhibition at Philadelphia in the same year. His demonstrations there, as in the case of his lectures, were confined to his receiver type instruments. Though he had a model of the liquid transmitter there, he did not use it.[11] Moreover the judges make no mention of any liquid transmitter in their official reports of Bell's exhibits, though they describe his receiver type instruments in detail.[12]

In the trial of the Dowd case, the only occasion when the claims of Bell and Gray came directly into conflict in court, at no time, during his entire nine days on the stand did Bell even mention his first transmission of speech on March 10, 1876, when a liquid transmitter was used.[13] In his patent of January 30, 1877,[14] on *Improvement in Electric Telegraphy*, when, if ever, he would be expected to consolidate all improvements in the telephone to which he was entitled, Bell reiterates his preference for receiver type transmitters but refers very guardedly to the possibility of other types, in these words:

> I have heretofore described and exhibited such other means of transmitting sound, as will be seen by reference to the proceedings of the American Academy of Arts and Sciences, Volume XII.

It should be noted that he merely "described and exhibited" another type of transmitter. There is no claim to having invented it, either in the foregoing extract, nor in the article to which it refers.

It seems clear that, at least during the first four years of the telephone, Bell did not feel justified in claiming the liquid transmitter. Not until after the settlement of the Dowd case in 1879, by compromise between the contending parties, one of the terms of which was a division of all telephone royalties between them during the life of the patents in suit,[15] did Bell begin to feel at all secure in his position with reference to the liquid transmitter. It was in 1880 that Bell made his first attempt to claim priority in conceiving it.[16]

Bell's skill as a tactician reaped its full reward. During the crucial period of the struggle for supremacy, Gray and his advisers were completely misled. Gray was present at Bell's Cen-

[7] *The Bell Telephone*, p. 379; *American Bell Telephone et al. vs People's Telephone Company*, I 324–325, II 1653.
[8] *The Bell Telephone*, pp. 95–96.
[9] A. Graham Bell, "Researches in Telephony," Proc. Am. Acad. Arts and Sci., Vol. XII, first paper.
[10] *The Bell Telephone*, p. 95.
[11] *Dowd Case Record*, I 494, 544; *United States Supreme Court Reports*, Vol. 126, p. 323; *The Bell Telephone*, p. 100.
[12] *Dowd Case Record*, I 167–171, II 31–35.
[13] *Dowd Case Record*, I 440–534.
[14] Patent No. 186,787 by A. G. Bell.
[15] Affidavit of George Gifford. Department of the Interior; in the Matter of the Petition for Leave to Bring a Suit in the Name of the United States to Cancel Patent No. 174,466, Issued March 7, 1876, to Alexander Graham Bell. American Bell Telephone Company. p. 16.
[16] *American Bell Telephone Company vs. People's Telephone Company*, 1880. Repeated in *The Bell Telephone*, pp. 83–88.

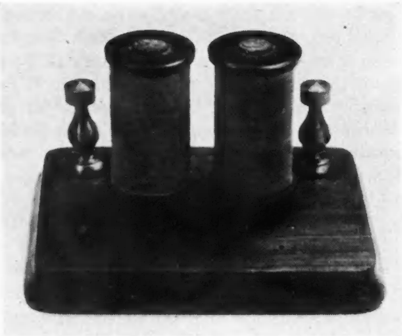


FIG. 6. Gray's first electromagnetic receiver.

tennial exhibition but did not see any liquid transmitter.[17] Had he seen it he would, of course, have instantly recognized it as his own and proceeded accordingly. Not knowing of its existence, he became convinced that he must concede priority to Bell because he supposed that the first successful performance of the telephone at Bell's hands had been with a transmitter quite different from his own. Gray summarized the case, in a letter written just before his death, in these words:

> It was not till eight or ten years—at least a long time after the telephone was in use—that I became convinced, chiefly through Bell's own testimony in the various suits, that I had shown him *how* to construct the telephone with which he obtained his first results.[18]

With this brief summary of the evidence that Bell made no claim to the microphone type of transmitter until four years after the time of his first experiment with it, we turn from consideration of the part played by the transmitter in the telephone contest to the part played by the receiver. Gray's priority is perhaps even more pronounced here than with the transmitter, for, as will be adduced presently, he had unchallengeable patent priority on this point. Legal precedence is not the point at present under emphasis, however. It is rather the relative actual accomplishment of the two men, which is often quite distinct from mere legal advantage. The fact in connection with this part of the telephone contest is that *Elisha Gray had made and publicly used several types of telephone receiver many months before Bell constructed his first one.*

There are many ways in which the fluctuating electric currents that carry telephone messages from one place to another can be converted back into sounds resembling the sounds that created them. But of all the ways that are possible, the one that has almost universally prevailed in practice is to cause the fluctuating currents to act upon a diaphragm, especially a metal diaphragm, through the agency of an electromagnet. Through all its variations, the telephone receiver has embodied this principle almost exclusively. Any inquiry, therefore, into precedence in development of the telephone receiver, as between Bell and Gray, must center on the question of which of the two first developed and used the combination of an electromagnet and

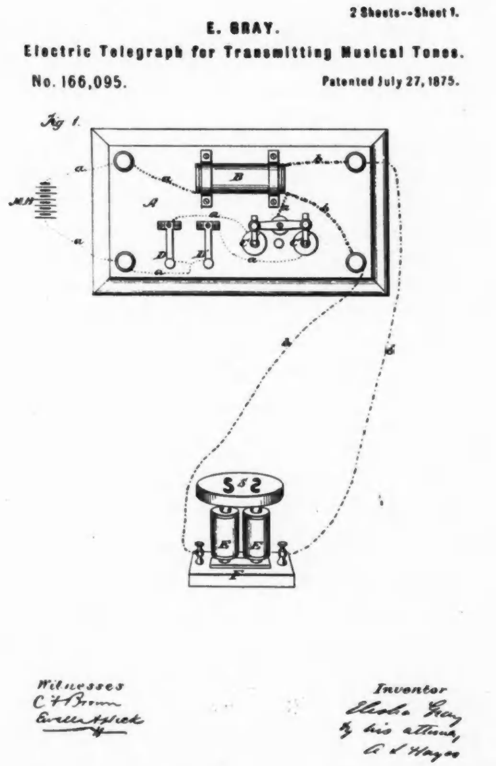


FIG. 7. Gray's first receiver, connected to reproduce the musical tone from a buzzer; taken from his American patent of 1875.

[17] *Dowd Case Record*, I 126–127.
[18] Letter from Gray, published in *The Electrical World and Engineer*, Feb. 2, 1901, p. 199.

FIG. 8. Gray's second receiver.

diaphragm, especially an electromagnet and metal diaphragm, to produce sound electrically.

Aside from Bell's use of one of his unmodified harmonic telegraph instruments as a makeshift receiver, his first attempt to construct an electromagnetic telephone receiver was in the summer of 1875, in an episode which is justly considered to be nearly as significant as his first experiment with the prototype of the modern transmitter in 1876. The diaphragm of this first receiver was of parchment however, not metal. This is the instrument of Fig. 2. The first occasion upon which Bell publicly demonstrated a receiver possessing a metal diaphragm was the Centennial exhibition in Philadelphia in June, 1876, a year later.

But while it is impossible to attribute to Bell the development of the telephone receiver earlier than June 1875, Elisha Gray had constructed and publicly demonstrated, as early as 1874 and before February 1875, not less than four receivers which were prototypes of the modern telephone receiver. Each one of these consisted of a pair of electromagnets, near the end of which was placed a diaphragm of some form. The receivers differed from each other primarily in the variations in this diaphragm.

Gray's first electromagnetic receiver is shown in Fig. 6. This instrument was given to Oberlin College by Elisha Gray personally, some time during the eighties, at which time he made a statement of its history. The diaphragm, originally an ordinary shoe-blacking box supported near the poles of the electromagnet, has been lost sometime during the intervening years. The way the instrument was used is shown in Fig. 7. It was made in May 1874 and was used during the same month in demonstrations in Boston, New York and Washington before a number of officials and directors of telegraph companies and in the presence of representatives of the Smithsonian Institution and the United States Naval Observatory.[19] This receiver was patented in England in July 1874 and in the United States in July 1875, nineteen months and seven months respectively before Bell's patent of 1876.

Gray's second receiver (Fig. 8) was made in Chicago in July 1874 and shown in operation in the same month and in the following month. Gray took it to England with him that same autumn and publicly demonstrated it there, showing it, among others, to Professor Tyndall at the Royal Institution in London.[20] A number of these receivers were made, two of which are in existence today. The original model is in the possession of the Smithsonian Institution.

Gray's third receiver was the progenitor of the modern loudspeaker. The receiver proper, including a metal diaphragm, was mounted upon a wooden sounding box which acted as a resonator (Fig. 9). The original is in possession of the Smithsonian Institution. This receiver was made in December 1874, and used in the same month in a public demonstration in the Presbyterian Church at Highland Park, Illinois.[21]

Gray's fourth receiver constituted another attempt to utilize resonance as a means of amplifying sound, but in a smaller and more sensitive instrument, capable of use on longer lines than the wooden sounding box type was designed for. It was used for the first time in February 1875 on a line between Chicago and Milwaukee.[22] It is illustrated in Fig. 10, taken from the record of the Dowd case. The original model has not survived. This is regrettable,

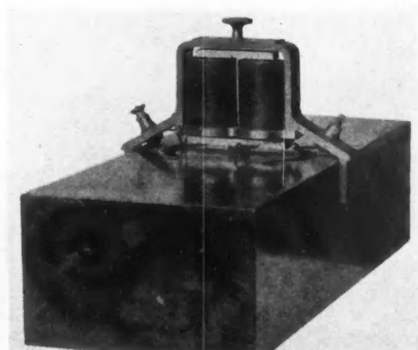


FIG. 9. Gray's third receiver.

[19] *Dowd Case Record*, I 74, 119.
[20] *Dowd Case Record*, I 74–76, 81, 114–116, 155–156.
[21] *Dowd Case Record*, I 86.
[22] *Dowd Case Record*, I 87–88, 122, II 770.

since it came closer to the conventional form of telephone receiver as it later developed than did any of the previous forms.

It is evident, from the documentation on the foregoing receivers, that, during the year preceding the earliest date that Bell can claim, Gray had made and exhibited to some hundreds of witnesses who were qualified to comprehend their principle and importance, four types of telephone receiver, all of which possessed thin metal diaphragms and hence anticipated the design of the modern telephone receiver much more closely than did the receiver which Bell first devised in 1875. But even these four were not his first successful ventures in this field. He had made, used, and published descriptions of at least three others prior to any of these four.[23] But since the earlier models acted on a principle which could not have produced the modern receiver, they are not included here.

It should be understood that the tests and demonstrations made with these receivers during 1874 and 1875 involved only the production of musical tones from intermittent currents produced by tuned buzzers in the circuits. This was not because the receivers were incapable of reproducing speech, for three out of the four were capable of reproducing speech excellently, as later tests showed.[24] That Gray had a strong suspicion that his receivers possessed this ability is made clear by his testimony before the Patent Commissioner,[25] but it was not possible for him to test their ability to reproduce speech because at that time no method was known for getting speech *into* an electric system to be so reproduced. This had to await the invention of the transmitter.

[23] *Speaking Telephone Interferences, Elisha Gray's Case.* United States Patent Office, 1880, pp. 6, 9, 10, 42, 61, 62, 63; *Dowd Case Record,* I 72, 73, 74, II 678, 763, 765; G. B. Prescott, *The Speaking Telephone* (Appleton, ed. 1, 1878), pp. 155, 188, 193; *Scientific American Supplement,* Feb. 5, 1876, p. 92; Gray's patents No. 166,096 of July 27, 1875 and No. 210,776 of Dec. 10, 1878.

[24] *Dowd Case Record,* I 92–94, 180, 183, 229, 261.

[25] *United States Patent Office,* 1888. Gray's "Petition before Commissioner of Patents," pp. 44–45.

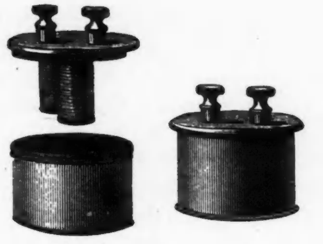


FIG. 10. Gray's fourth receiver.

For some reason, Gray's priority in the development of the telephone *receiver* has been allowed to escape general notice. That story has now been told. As for the *transmitter*, the story of how he was maneuvered by Bell and Bell's associates out of the maintenance of his proper claim to priority in design of the microphone type of telephone transmitter has also now been told. Gray's loss of credit for these two major contributions to the development of the telephone was quite possibly due in part to his own ineptitude as a tactician. As to the facts of his priority in both fields there is little room for controversy. The curious popular interpretation of those facts which gives exclusive credit to Bell for designs previously recorded and devices previously constructed by Gray is, alas, all too typical of common practice throughout the whole history of electrical communication.[26]

After Gray's death, a scrap of paper was found among his belongings,[27] on which he had written in his quaint scrawl, during some moment of profound discouragement, the following epitaph on the telephone contest:

> *The history of the telephone will never be fully written. It is partly hidden away in 20 or 30 thousand pages of testimony and partly lying on the hearts and consciences of a few whose lips are Sealed,—Some in death and others by a golden clasp whose grip is even tighter.*

[26] For an apt and informed statement on the pronounced trend in this direction during the entire history of electrical communication see *Old Wires and New Waves,* by Alvin Harlow (D. Appleton-Century Co., 1936), p. 162.

[27] Original autograph copy in the author's possession.

Advanced Laboratory Experiments in Acoustics, Including a New Method for Measuring the Absorption of Sound in Tubes

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SINCE it is the demand for high quality microphones, loudspeakers and sound reproducing systems that has provided the impetus for the great development of acoustical science in recent years, it is quite natural to find much space devoted to the theory of these devices in modern textbooks on sound. For the same reason it is desirable to give students an opportunity to experiment with electro-acoustic apparatus in the laboratory so that they can master the theory by applying it to problems. Inasmuch as most equipment for quantitative measurements in acoustics is elaborate and expensive, it seems worth while to publish a description of an apparatus which can be constructed quite cheaply, but which nevertheless can be used for a great variety of instructive experiments occupying half a dozen or more laboratory periods in a graduate or advanced undergraduate course. The quantities which can be measured are acoustic impedance, the effective mass, stiffness and damping of diaphragms, absorption coefficients, and the absorption of sound in tubes. The apparatus can be used to generate sound waves of known phase and intensity in a tube. It also provides a striking, elementary lecture-demonstration experiment which may be described as an electrical Kundt's tube, the points of resonance being shown by the rise and fall of a voltmeter, as a plunger is moved along the tube; this particular experiment, because of its more general interest, is described in less technical language at the end of the present article.

In brief, the procedure is to replace the horn of a moving coil horn-type loudspeaker with a uniform tube closed by a tight-fitting movable piston. The electrical impedance of the coil is then measured in an a.c. bridge at different tube lengths, yielding an approximately circular impedance locus. From this locus one can obtain the corresponding mechanical impedances by a graphical construction, and thence all data necessary to calibrate the unit. Finally, the impedance is measured with the horn replaced, and one can then calculate the acoustic impedance of the horn, the sound pressure in the throat given in magnitude and phase in terms of the applied voltage, the acoustic power output, the mechanical and electrical efficiency, and the mechanical impedance of the diaphragm. If, instead of replacing the horn, a plug of absorbing material is placed in the tube, its absorption coefficient can be measured.

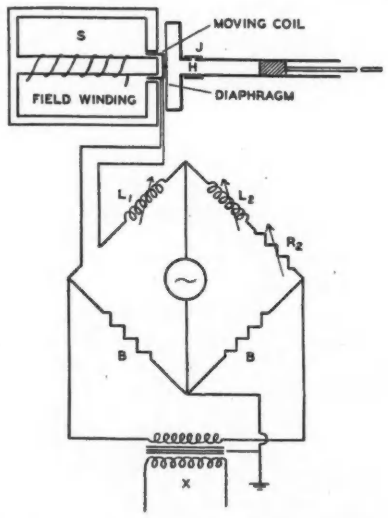

Fig. 1. Schematic diagram of apparatus.

Description of Apparatus

In Fig. 1, S may be either a commercial horn-type unit, or the magnet of a cone-type speaker with parts screwed on the front plate to permit the mounting of a metallic diaphragm. The latter may be either plane sheet metal or pressed Duralumin,[1] depending on considerations which will be mentioned later. A snugly fitting sleeve J covers the hole H when measurements are being made, and can be slid back to permit the passage of air when the piston is moved.

[1] Pressed Duralumin diaphragms with coils ready mounted may be obtained from Fox Sound Equipment Corp., 3120 Monroe St., Toledo, Ohio, or from Racon Electric Co., 52 East 19th St., New York City.

Inasmuch as the bridge has a 1 to 1 ratio and quite low impedance arms, it can be wired up using variometers and resistance boxes without special shielding. A symmetrically wound astatic transformer is at X. The most convenient a.c. source is a beat-frequency oscillator of good frequency stability and wave form. If this is not available, a complete set of measurements can be carried out at a single frequency but then, of course, one cannot study the frequency dependence of the measured quantities. If the speaker magnet is used as a filter choke, the 120-cycle ripple induced in the moving coil makes it difficult to balance the bridge. This voltage can be balanced out by connecting in series with the moving coil, a fixed coil of suitable number of turns wound around the center leg of the magnet.

Procedure and Calculations

We will leave for a later section the discussion of sound absorption by the tube, and first describe the measurement of acoustic impedance and other loudspeaker characteristics upon which it has negligible effect.

Observations. The theory of the method has been worked out in detail and presented in earlier papers,[2, 3] but for the sake of completeness the experimental procedure and the principal equations will be given here very briefly without proof. It is not expected that the reader will be able to understand the reasons for the procedure outlined in these sections without referring to the earlier papers.

First demagnetize the field magnet or clamp the diaphragm so as to measure the impedance of the moving coil when it is not vibrating. This impedance is represented by OO' in Fig. 2. With the field excited and the diaphragm free to vibrate, leave the bridge reactances fixed, but vary the resistance and the tube length to locate a succession of points of minimum motional impedance α, β, γ. If these occur at tube lengths l_α, l_β, l_γ, then $l_\beta - l_\alpha = l_\gamma - l_\beta = \frac{1}{2}\lambda$. The first point comes at a tube length somewhat less than $\frac{1}{2}\lambda$ and we will set the difference $\frac{1}{2}\lambda - l_\alpha = r$. Next

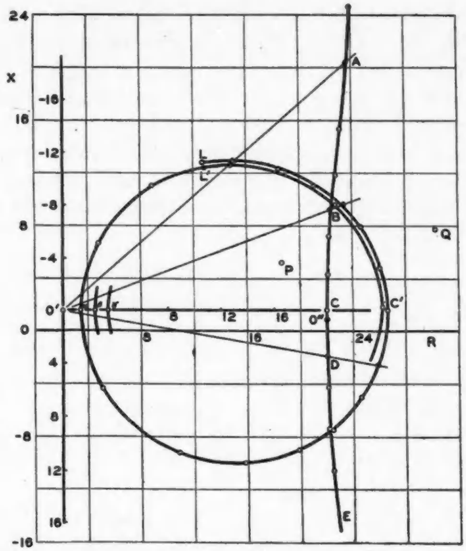

Fig. 2. Data obtained with plane phosphor bronze diaphragm.

measure the impedance at a number of tube lengths between 0 and l_α and plot them on rectangular coordinates to obtain the approximately circular locus of motional impedances. It is not necessary to make as many observations as are shown in Fig. 2, and there will be no need for points beyond l_α except the two which have already been observed. Finally remove the tube, replace it with the unknown acoustic load (for example, the speaker horn), and measure the electrical impedance once more (point P, Fig. 2).

Calibration. Each motional impedance Z_{EL} is related to the corresponding mechanical impedance Z_M by the equation[4]

$$Z_{EL} Z_M = k^2, \tag{1}$$

where k is a constant depending on the dimensions of the moving coil and the strength of the magnetic field. Hence to construct the locus $ABCDE$ of mechanical impedances, draw from O' a line through each point of electrical impedance and mark off on each line a distance inversely proportional to the distance from O' to the point. The constant of proportionality can be chosen at will to place the locus conveniently on the graph. Use the same construction to locate

[2] R. D. Fay and W. M. Hall, "The Determination of the Acoustical Output of a Telephone Receiver from Input Measurements," J. Acous. Soc. Am. 5, 46 (1933).

[3] C. K. Stedman, "A New Treatment of the Horn-Diaphragm Coupling Chamber for Receiver Measurements," J. Acous. Soc. Am. 7, 265 (1936).

[4] This k can easily be distinguished from the other value, $2\pi/\lambda$, which follows, by the context.

the point Q in line with P. It can be shown that the vectors $O'A$, $O'B$, etc. are the mechanical impedances at the diaphragm, although it remains of course to establish their scale in mechanical ohms, and to find what portion of the total impedance is due to the diaphragm and what portion to the acoustic load. This can readily be done because the acoustic part can be calculated from the equation[5]

$$X_A = -(A_1^2/A_2)\rho c[\cos^2 kr \cot k(l+r) + \sin kr \cos kr], \quad (2)$$

where X_A is the air load reactance, pressure $\times$ area/particle velocity, at the diaphragm; A_1, the diaphragm area; A_2, the sectional area of the tube; ρ, the density of air; c, the velocity of sound in air; l, the tube length; $r=\frac{1}{2}\lambda - l_a$; and $k=2\pi/\lambda$.

If we may assume that the smallest air-load resistance is negligible in comparison with the diaphragm resistance, the origin O'' of air-load impedances must lie somewhere on a vertical line through C. Its position is most accurately found by plotting a curve of observed mechanical reactances CA, CB, etc. against corresponding values calculated from Eq. (2). The result should be a straight line of which the intercept where $X_A=0$ fixes the location of O'' and whose slope establishes the scale of mechanical impedances in mechanical ohms per centimeter from the origin O'. Mechanical reactances above the axis are negative.

Acoustic impedance of horn. The real part R_{AL} and the imaginary part X_{AL} of $O''Q$, the air load impedance due to the horn, can now be read from the graph in mechanical ohms. The acoustic impedance $M+jN$ is given by

$$M = \frac{R_{AL}\{(A_1/A_2)\rho c \cot kr\}^2}{R_{AL}^2 + \{X_{AL} + (A_1^2/A_2)\rho c \cot kr\}^2}, \quad (3)$$

$$N = \frac{\rho c \cot kr}{A_2} \times \frac{R_{AL}^2 + X_{AL}\{X_{AL} + (A_1^2/A_2)\rho c \cot kr\}}{R_{AL}^2 + \{X_{AL} + (A_1^2/A_2)\rho c \cot kr\}^2}. \quad (4)$$

Diaphragm impedance. For simplicity we have thus far treated the problem as though the diaphragm moved like a plane piston of area A_1. Such is not the case, however, and since the average diaphragm velocity is less than the coil velocity, the diaphragm behaves as a mechanical transformer of ratio k' where

$$k' = \frac{\text{volumetric rate of air displacement}}{\text{velocity of coil} \times A_1}.$$

It follows that the impedances measured from O'' are not truly load impedances measured at the moving coil, but are really mechanical impedances of the air column at the diaphragm which are greater by a factor $1/k'^2$, and that the mechanical impedance Z_D of the diaphragm is not $O'O''$ but

$$Z_D = k'^2 O'O''. \quad (5)$$

To obtain the value of k' note that, from Eq. (1) and Fig. 2, $Z_M Z_{EL} = (k'^2 O'C)(O'C') = k^2$, or

$$k^2/k'^2 = (O'C)(O'C'). \quad (6)$$

Furthermore it can be shown that if F is the force in dynes on the moving coil due to a current of i amp., $F=ki(10^7)^{\frac{1}{2}}$. So k can be either measured or calculated and substituted in Eq. (6) to give k'. The value of Z_D is then obtained from Eq. (5).

Sound pressure and power output. It is proven in reference 3 that the vector excess pressure p in the throat of the horn is given by

$$[pA_1/E] = [(Z_{AL} Z_{EL}(k'/k)(10^7)^{\frac{1}{2}})/Z_{ET}] \quad (7)$$

where E is the vector applied voltage, $Z_{AL}=O''Q$ mechanical ohms, $Z_{EL}=O'P$ electrical ohms, and $Z_{ET}=OP$ electrical ohms. By analogy with the corresponding electrical equation the power is

$$\left(\frac{1}{E^2}\right)\left(\frac{p^2 A_1^2 R_{AL}}{R_{AL}^2 + X_{AL}^2}\right) \times 10^{-7} \text{ watts/volt}^2,$$

or $$\left(\frac{1}{E^2}\right)\left(\frac{p^2 M}{M^2+N^2}\right) \times 10^{-7} \text{ watts/volt}^2.$$

Efficiency. The electrical efficiency is the ratio of the real parts of $O'P$ and OP. The mechanical efficiency is the ratio of the real parts of $O''Q$ and $O'Q$.

Measurement of absorption coefficient. Absorption coefficients can be measured for normal incidence by placing a plug of the material in the

[5] Reference 3, Eq. (5).

tube, and measuring the resulting air-load impedance at the diaphragm in the manner already described. It is left as a problem for the student to derive an expression for the specific acoustic impedance and absorption coefficient of the material in terms of the measured air load, and to decide upon the most suitable tube length.

Absorption of Sound in Tubes

Theory. In the foregoing paragraphs and in the earlier articles no account was taken of the effect of absorption of sound by the tube. The influence of dissipation is very evident in Fig. 2, but nevertheless its effect on the measurements described thus far is not appreciable. On the other hand, under suitable conditions the effect becomes relatively much greater, and the attenuation factor of the tube can be measured with an accuracy comparable with that attained using far more complicated apparatus.[6] Since it has not been published elsewhere, this part of the theory will be given in somewhat more detail than the rest.

The general equations for pressure p and volume current V at a distance x along a tube are[7]

$$V = V_1 \cosh \alpha x - (p_1/Z_L) \sinh \alpha x, \tag{8}$$

$$p = p_1 \cosh \alpha x - V_1 Z_L \sinh \alpha x, \tag{9}$$

where p_1 is the sound pressure and V_1 the volume current at $x=0$; and α is the propagation constant of the tube, and Z_L its characteristic acoustic impedance. Also,

$$\alpha = a + jb = \left(\frac{\gamma'}{Rc}\right)(\tfrac{1}{2}\omega)^{\frac{1}{2}} + j\left(\frac{\omega}{c}\right)\left[1 + \left(\frac{\gamma'}{R}\right)\left(\frac{1}{2\omega}\right)^{\frac{1}{2}}\right] \tag{10}$$

and $Z_L = (P_0\gamma\rho)^{\frac{1}{2}}/A_2 = \rho c/A_2$ approximately, where R is the radius of the tube, μ the kinematic coefficient of viscosity of air, γ the ratio of specific heats, ν the coefficient of thermal diffusivity, P_0 the average atmospheric pressure, $\gamma' = \mu^{\frac{1}{2}} + (\gamma^{\frac{1}{2}} - 1/\gamma^{\frac{1}{2}})\nu^{\frac{1}{2}}$, and ρ, ω, c, and j have their customary significances.

The volume current is zero at the end of a closed tube; hence, from Eq. (8),

$$0 = V_1 \cosh \alpha l - (A_2 p_1/\rho c) \sinh \alpha l$$

or, regarding the tube itself now as an unknown acoustic load,

$$p_1/V_1 = M + jN = (\rho c/A_2) \coth \alpha l.$$

To obtain the resulting acoustic impedance at the diaphragm, Z_α, this must be combined in parallel with the acoustic capacitance of the horn-diaphragm coupling chamber, giving

$$Z_\alpha = \frac{(-j/\omega C)(\rho c/A_2) \coth (a+jb)l}{(-j/\omega C) + (\rho c/A_2) \coth (a+jb)l}.$$

On expanding into exponentials, separating real and imaginary parts, and for brevity substituting[8] $A_2/\rho c\omega C = n$, this becomes

$$Z_\alpha\left(\frac{\rho c}{A_2}\right)\left[\frac{\dfrac{n^2}{2}\left(\dfrac{e^{4al}-1}{e^{2al}}\right) - jn\left(\dfrac{e^{4al}+1}{2e^{2al}} + n \sin 2bl + \cos 2bl\right)}{\dfrac{n^2+1}{2}\,\dfrac{e^{4al}+1}{e^{2al}} + (1-n^2)\cos 2bl + 2n \sin 2bl}\right].$$

But

$$\frac{e^{4al}-1}{e^{2al}} = e^{2al} - e^{-2al} = 1 + 2al + \frac{4a^2l^2}{2} + \cdots$$
$$-1 + 2al - \frac{4a^2l^2}{2} + \cdots \doteq 4al,$$

$$\frac{e^{4al}+1}{e^{2al}} = e^{2al} + e^{-2al} = 1 + 2al + \frac{4a^2l^2}{2} + \cdots$$
$$1 - 2al + \frac{4a^2l^2}{2} + \cdots \doteq 2 + 4a^2l^2;$$

therefore

$$Z_\alpha = \left(\frac{\rho c}{A_2}\right)\left[\frac{2aln^2 - jn(1 + 2a^2l^2 + n \sin 2bl + \cos 2bl)}{(n^2+1)(1+2a^2l^2) + (1-n^2)\cos 2bl + 2n \sin 2bl}\right]. \tag{11}$$

In order to give some insight into the very curious behavior of this function, it has been computed over a range of four half-wave-lengths for a tube 1.25 cm in diameter driven at 975 cycle/sec., corresponding to the conditions in Fig. 2. The results are plotted in Figs. 3 to 6. In Figs. 3 and 4 the resistance and reactance are separately plotted as functions of the angle $2bl$. The four cycles have been superimposed so as to make clear the increasing influence of damping as the tube grows longer. In Fig. 5 they are plotted on rectangular coordinates as an acoustic impedance vector diagram. Finally, in Fig. 6 is shown the region near the origin plotted on a magnified scale, and the complete vector diagram of electrical impedances constructed graphically by the reverse of the method already described. In drawing Fig. 6, $O'O''$ was made equal to the previously measured diaphragm resistance and the diaphragm was supposed tuned to resonance, that is, zero diaphragm reactance. In addition to the lowering and broadening of the peaks in Fig. 4 their increasing dissymmetry is to be noticed, a dissymmetry which reveals itself in Fig. 5 as a nearly constant depression of the centers of the (approximate) circles. A further important consequence is the fact that neither the minimum absolute

[6] W. P. Mason, "Propagation Characteristics of Sound Tubes and Acoustic Filters," Bell System Reprint No. B-295, or Phys. Rev. **31**, 283 (1928).

[7] These equations are taken with a slight change of notation from W. P. Mason, "A Study of the Regular Combination of Acoustic Elements, with Applications to Recurrent Acoustic Filters, Tapered Acoustic Filters, and Horns," Bell System Tech. J. **6**, 261 (1927).

[8] It is proven in reference 3 that $A_2/(\rho c\omega C) = \cot kr$, a quantity which can be determined empirically even when the wave-length is so short that the concept of acoustic capacitance of the chamber breaks down.

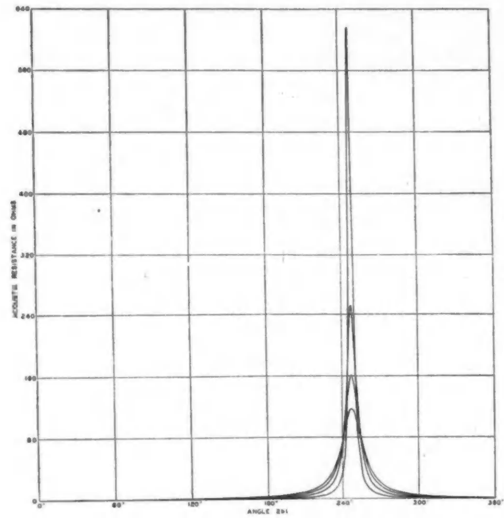

FIG. 3. Theoretical curve of acoustic resistance of closed tube in parallel with coupling chamber.

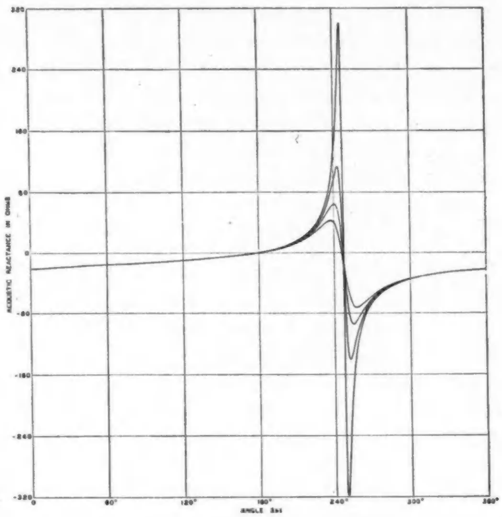

FIG. 4. Theoretical curve of acoustic reactance of closed tube in parallel with coupling chamber.

value of Z_α nor the point where the reactance is zero, occurs when the resistance is a minimum.

We must now remember that experimentally the procedure will be the reverse of that which we have just carried out. The electrical impedance locus will first be obtained, from it the mechanical impedance locus will be constructed, and our purpose will be to infer the value of a. To this end it has been found that the quantity most easily determined experimentally and computed from Eq. (11) in terms of a is the value of the resistance when the reactance equals zero. These are the distances $O''A$, $O''B$, etc. in Fig. 6.

From Eq. (11) the reactance is zero when

$$1+2a^2l^2+n\sin 2bl+\cos 2bl=0;$$

or, substituting for n,

$$1+2a^2l^2+\cot br\sin 2bl+\cos 2bl=0,$$

or

$$(1+2a^2l^2)\sin br=-\sin b(l+2r).$$

If a is zero the roots are

$$\begin{cases} 2bl=m\pi, & m=1, 3, 5, \cdots, \\ 2bl=2\pi m-2br, & m=1, 2, 3, \cdots; \end{cases}$$

that is

$$\begin{cases} l=m\lambda/4, & m=1, 3, 5, \cdots, \\ l=m\tfrac{1}{2}\lambda-r, & m=1, 2, 3, \cdots. \end{cases}$$

The first set of roots contains those which we require. Since $2a^2l^2$ is a small quantity the roots will be shifted only slightly; in fact the points corresponding to $2bl=m\pi$ lie on the dotted line $O''P$ in Fig. 6. Therefore we substitute $2bl=m\pi+\epsilon$, $m=1, 3, 5, \cdots$. Then $1+2a^2l^2-n\epsilon-1=0$ or

$$\epsilon=2a^2l^2/n. \tag{12}$$

Inserting $2bl=m\pi+(2a^2l^2/n)$ in the real part of Eq. (11), we get

$$R_\alpha=(\rho c/A_2)al/[1-((1-n^2)/n^2)a^2l^2]. \tag{13}$$

Unless the tube is very long and narrow the correction term in the denominator can be neglected. In the example given below it amounts to less than 1 percent so that R_α reduces to a very simple expression,

$$R_\alpha=(\rho c/A_2)al,\quad l=\lambda/4,\quad 3\lambda/4,\quad \cdots. \tag{14}$$

Choice of receiver for measurement of R_α. A glance at Fig. 2 will convince the reader that the load due to the acoustic resistance of the tube is so small in comparison with the diaphragm resistance, that R_α could not be measured accurately. To overcome this difficulty a number of courses are open to the experimenter: (1) the value of k [Eq. (1)] may be increased by strengthening the field or increasing the length of wire in the moving coil; (2) a diaphragm may be found that has less damping; (3) A_2 may be decreased, although this possibility is of little interest if one desires to measure a for a given tube; (4) the diaphragm area A_1 may be increased; (5) the value of k' [Eq. (5)], which determines the effective diaphragm area, could be increased by using a diaphragm that moves more nearly as a piston.

In obtaining data for Fig. 2 a plane phosphor bronze diaphragm 4.6 cm in diameter and 0.054 cm thick was used, whereas for Fig. 7 the data were obtained with a pressed Duralumin diaphragm similar to the type designed by Wente

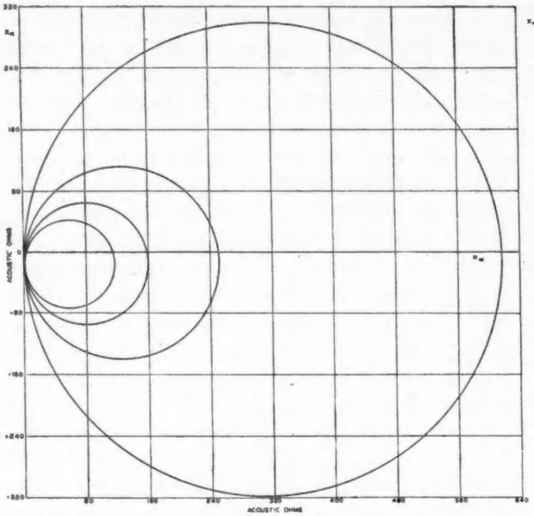

FIG. 5. Acoustic impedance diagram of closed tube in parallel with coupling chamber.

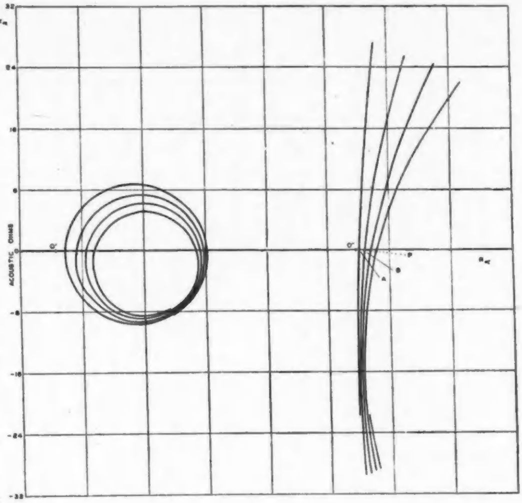

FIG. 6. Computed locus of electrical impedances.

and Thuras.[9] By chance, k and the damping are not greatly different in the two diaphragms, but the values of k'^2 are roughly 0.09 and 0.6, and those of A_1^2 are 1400 and 280 cm^4, respectively. Hence the scale of air loads relative to the diaphragm impedance is multiplied by a factor $A_1^2k'^2=33$, and the small air load due to R_α is brought into prominence.

Inasmuch as large mechanical impedances lead to small motional impedances, such a receiver would not be satisfactory for measuring the acoustic impedance of horns unless the throat area A_2 was so large as to reduce the acoustic impedance to 7 or 8 ohms. Even then to make accurate measurements it would be necessary to raise the efficiency by increasing the field strength. (The present field is somewhat less than 10,000 gauss.) Another very important characteristic of diaphragms is their effective mass and stiffness. A massive plane diaphragm such as the one just described must be operated within a few cycles of resonance lest its reactance completely mask the air load. The Duralumin diaphragm, on the other hand, is far lighter and is useful for measurements over a wide range of frequencies.

Numerical example. For the data plotted in Fig. 7, a tube 1.26 cm in diameter was driven at 575 cycle/sec. The observations are made in just the same way as for Fig. 2, but two more points X and Y are observed. This is done by leaving the variometers fixed and varying the resistance and tube length to obtain balance just as was done for α, β, γ in Fig. 2. (These points are coincident at O' in the present case.) The tube length is $l=\lambda/4$ at Z, $l=3\lambda/4$ at Y, and $l=5\lambda/4$ at X. The corresponding mechanical impedances are $O''C$, $O''A$, $O''B$. The scale is established as before by plotting observed reactances along MM' against reactances calculated from Eq. (2), and gives 0.752 acoustic ohm/div. (or $0.752A_1^2$ mechanical ohm/div.). To locate the origin O'', the reactance of $O''C$ is given by the same graph that establishes the scale, and the resistance of $O''C$ is obviously $CA/2$, since the tube length is $\lambda/4$ at C and $3\lambda/4$ at A. Strictly speaking, the resistances CA, CB should be measured in line with O'', but with the diaphragm tuned nearly to resonance the error is negligible. The advantage of tuning the diaphragm lies in the fact that only the two points X, Y need be observed instead of a whole series of points around the inner loci of electrical impedance.

By measurement on the graph, $CA=AB=1.45$ div., which corresponds to 1.09 acoustic ohms. The difference in tube length is $\lambda/2=29.40$ cm. Hence, from Eq. (14), $a=0.00113$ cm^{-1}.

To compute a from Eq. (10) we have $a=(\gamma'/Rc)(\frac{1}{2}\omega)^{\frac{1}{2}}$, $\rho\mu=0.000,182-0.000,000,493(23°-t)$ poises at t°C, $\gamma=1.405$, $\nu=1.91\mu$.[10] Substitution of the proper values for ω and R gives $a=0.00113$ cm^{-1}. The observed value also agrees closely with that obtained by interpolation from Mason's curves,[6] namely 0.00115 cm^{-1}.

[9] E. C. Wente and A. L. Thuras, "A High Efficiency Receiver for Horn Type Loudspeakers of Large Power Capacity," Bell System Tech. J. 7, 140 (1928).

[10] In Loeb, *Kinetic Theory of Gases*, ed. 1, p. 215, the coefficient of thermal conductivity ν' is given by $\nu'=1.91C_v\mu'$ where μ' is the coefficient of viscosity. Hence $\nu=\nu'/\rho C_v=1.91\mu'/\rho=1.91\mu$.

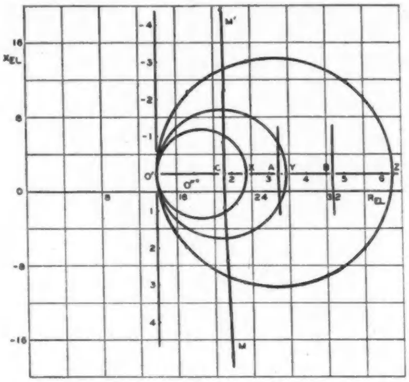

FIG. 7. Data obtained with Duralumin diaphragm.

Velocity of sound in tubes. By leaving the bridge reactances fixed and varying the tube length to obtain a balance, the successive points of maximum motional resistance (Z, Y, X, Fig. 7) can be located very accurately. Hence if the frequency is known accurately the velocity of sound in the tube can be measured certainly within one part in a thousand. Although the velocity does not depend on tube length, the measured half-wave-lengths grow longer as the tube-length increases. The correction, in case it is appreciable, can be calculated from Eq. (12).

Electrical Kundt's Tube

If, in place of the bridge circuit in Fig. 1, one simply connects the oscillator directly to the moving coil through a resistance of several hundred ohms, the current will remain approximately constant as the plunger is moved along the tube. Consequently if an a.c. voltmeter is connected across the coil its reading will rise and fall with the electrical impedance. A typical curve is shown in Fig. 8. Since the experiment is only qualitative, the hole H can be left slightly open to permit the plunger to be moved readily back and forth. The experiment is quite striking for a lecture-demonstration if one uses a large station voltmeter in conjunction with a copper oxide rectifier. Furthermore, it is possible to give a qualitative explanation of what takes place that can be understood by anyone with a knowledge of elementary acoustics and electricity.

The moving coil can be likened to the armature of a separately excited d.c. motor, for the applied voltage is equal to the sum of the IR drop and the back e.m.f. due to motion of the conductors in the magnetic field, while damping in the diaphragm corresponds to bearing friction and the air column to the useful load. When we move the plunger along the tube we are raising and lowering the load between a large maximum and a small minimum value, the current and the force being meanwhile held constant. It follows that when the load is large, the coil and diaphragm will move with small speed like a heavily loaded motor, and vice versa. Furthermore, when the coil moves slowly it can generate only a small back e.m.f., and the entire voltage across the coil will be small as at A, Fig. 8. When the load is smallest the speed, and hence also the voltage, is a maximum as at B.

The fact that the maxima of the curve, that is to say, the minimum values of air load, occur when the tube length is an odd multiple of $\lambda/4$ can be explained quite readily. When the tube length is $\lambda/4$, for example, a compression which had been caused by a forward motion of the diaphragm would travel to the end of the tube and return a half period later, just in time to cancel the rarefaction which would otherwise be caused by the backward motion of the diaphragm. Thus no matter how large its velocity, in the ideal case of no dissipation the diaphragm is unable to produce the slightest sound pressure in the end of the tube; in other words, it is a pressure node, and the air load is zero. At the

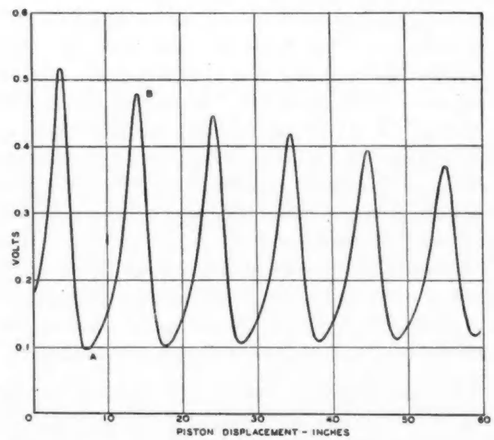

FIG. 8. Voltage across moving coil at constant current.

face of the plunger there is, of course, a pressure loop. By a similar argument it can be shown that when the tube length is a multiple of $\lambda/2$ there is a pressure loop at both ends and the air load is infinite. In the practical case the returning waves are slightly weakened by their journey along the tube, and so cannot quite cancel the new pulse at the diaphragm; thus the air load has a definite minimum greater than zero, and a finite maximum.

At the points of maximum load, the diaphragm coupling chamber behaves like an addition to the length of the tube, but it has no effect at the points of minimum load. Hence the dips in Fig. 8 do not lie halfway between the peaks but are all shifted a constant amount to the left. If the diaphragm is not driven at its resonant frequency both peaks and dips are shifted, but the distance between successive peaks or between successive dips is always $\lambda/2$.

When this experiment is performed it is most interesting to observe that the loudest sound is

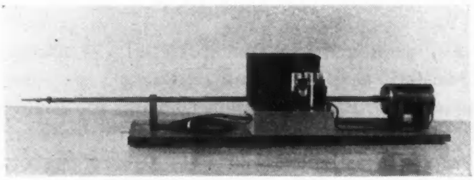

FIG. 9. Speaker unit and tuning fork oscillator.

emitted when the tube length is approximately an odd multiple of $\lambda/4$, and the weakest when it is a multiple of $\lambda/2$, just the reverse of what happens with the conventional Kundt's tube. This is a consequence of the fact that we are driving the diaphragm with an approximately constant alternating force, so that the greatest power output is drawn by the smallest air load, just as in the case of constant voltage electric power lines. In the Kundt's tube it is the diaphragm amplitude which is more nearly held constant, so that the greatest power is drawn by the largest load.[11]

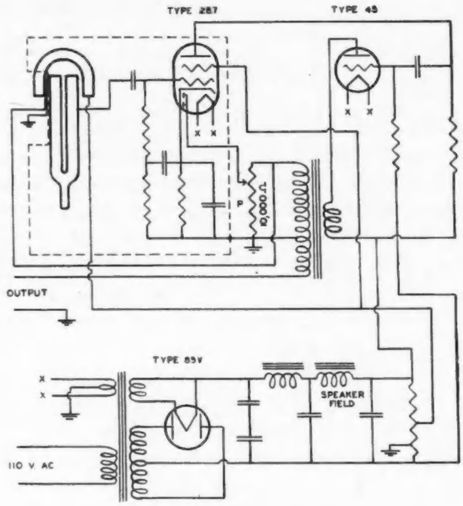

FIG. 10. Circuit of tuning fork oscillator.

Figure 9 shows a self-contained unit built for the Kundt's tube demonstration and for student laboratory measurements at a single frequency. The a.c. source is a tuning fork oscillator (Fig. 10) having an aluminum fork with iron inserts. If a drive coil and pick-up coil mounted on the same magnet are used with an aluminum fork, direct inductive coupling may be troublesome. This has been eliminated, as has the damping of one pole piece, by the use of electrostatic pick-up. Another feature is the use of the 2B7 diode electrodes to provide a bias by rectifying part of the output voltage. As the amplitude of the fork builds up, the increasing bias reduces the amplification until a stable operating point is reached. In this way there is no danger of the amplitude being limited by overloading the output tube and the wave-form is improved. The potentiometer P, which controls the bias, provides a convenient control of the output of the oscillator.

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[11] Strictly speaking it is the total force, diaphragm damping plus air load, that is held constant. To make the situation clear we may draw upon another analogy and liken the constant force to the e.m.f. of a battery, the damping to its internal resistance R_i, and the air load to the external load resistance R_L. The power output is $(E^2/(R_i+R_L)^2)R_L$ and is a maximum when $R_i=R_L$. Now the minimum load is usually much less than the diaphragm damping, and the maximum load much greater, so that really either our maximum or minimum load might give the greater sound output, or they might be equal. It is easy to show that if $R_{L\max}R_{L\min}$ is equal to, greater than, or less than R_i^2, the power drawn by the greater load will be accordingly equal to, less than, or greater than that drawn by the smaller load. The maximum power and sound output will, of course, lie somewhere between.

On the Relation of Two Mean Free Paths

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THE mean free path of gas molecules originally meant the average value of a very large number of the distances traveled by the molecules between collisions. This original concept is still frequently used in qualitative considerations, especially if they are of an elementary character. However, in the discussion, by the kinetic theory, of viscosity, thermal conductivity, and other problems involving the mean free path, it was found convenient to replace the consideration of distances by that of speeds, and to take the mean free path as the ratio of the average speed to the average frequency of collisions. Since the mean free path thus obtained depends on the manner in which the averages are taken, there are several mean free paths. One might recall in this connection, for example, the difference between the mean free path used by Maxwell and the one used by Tait. Since the emphasis in the present discussion is on distances rather than on speeds, however, it will be convenient to return to the original concept and to take as the mean free path the value of l given by

$$l = L/k = (x_1 + x_2 \cdots x_k)/k, \tag{1}$$

where k is the number of collisions occurring in a rather long distance L traversed by a single molecule, and $x_1, x_2, \cdots$ are the individual free paths between successive collisions.

We next introduce a quantity b which, for brevity, we may call the *mean expectation path.* The value of b is obtained as follows. Select some point at random in the distance L. It should be emphasized that it makes no difference whether the point selected is at the junction of two values of x (i.e., at a collision), or is somewhere in between, removed by a perceptible fraction of x from the end points of this x. Call the distance that the molecule has to go to its next collision y. Now

$$b = \text{average of all values of } y. \tag{2}$$

The purpose of this discussion is to consider the relation between l and b.

The point was first raised by Korteweg,[1] who took $b = l/2$. Clausius,[2] on the other hand, removed the factor 2 and took $b = l$ without, however, giving a detailed quantitative proof. Boltzmann[3] discusses the matter briefly, but also gives no formal proof. He says that a trivial example, for which he takes the chance of throwing an ace with an ordinary die, will illustrate the matter better than a lengthy discussion. Jeans[4] refers briefly to Boltzmann's illustration, but adds nothing to it, and Loeb[5] does not discuss the matter.

The argument of Korteweg is simply this. Suppose that the point selected lies somewhere on x_1. Then the average value of all possible values of y in x_1 is surely $x_1/2$. Similarly the average value of all possible values of y in x_2 is $x_2/2$, etc., and hence we should have

$$b = \tfrac{1}{2} \text{ average value of } x = \tfrac{1}{2} l. \tag{3}$$

Let us now carry out in detail the analogy suggested by Boltzmann. We are to make throws with a single die. Every ace is the analogy of a collision and does not contribute to the distance. Every other throw is the analogy of a free path of a certain length, say a unit length or, better, an arbitrary small element of length a. Start throwing at any instant. It is quite irrelevant whether the preceding throw was an ace, a collision, or not. The chance of having a collision at once, thus traveling no distance at all after the time instant chosen, is obviously

$$\tfrac{1}{6} = 1 - (5/6) = 1 - p.$$

The chance of going one unit a is $5/6$ or p, and the chance of then making a collision is $(1-p)$; hence the chance of going just one unit a is $p(1-p)$, the chance of going two units, and then making a collision is $p^2(1-p)$, etc. Now by a general theorem which is frequently applied in the kinetic theory, the average path b to be

[1] Korteweg, Arch. Néerlandaises **12**, 241 (1877).
[2] Clausius, Ann. d. Physik **10**, 92 (1880). See also the appendix to Vol. 3 of *Mechanische Wärmetheorie.*
[3] Boltzmann, *Gastheorie* (ed. 3, 1923), Vol. 1, p. 71, etc.
[4] Jeans, *Dynamical Theory of Gases* (1904), p. 246.
[5] Loeb, *Kinetic Theory of Gases.*

expected is obtained by multiplying any path y to be expected by the probability of occurrence of this y, and adding for all possible values of y. Thus

$$b=0\cdot a(1-p)+ap(1-p)+2ap^2(1-p)+\cdots$$
$$=a(1-p)(p+2p^2+{}^3p^3+\cdots+np^n)$$
$$=a(1-p)\left(\frac{p-p^{n+1}}{(1-p)^2}-\frac{np^{n+1}}{1-p}\right),$$

which, for values of p less than unity and for sufficiently large values of n, gives

$$b=a(1-p)\frac{p}{(1-p)^2}=\frac{ap}{(1-p)}.$$

Substituting the value 5/6 for p, we get $b=5a$.

Now in a very large number N of throws there will be approximately $N/6$ aces and $(5/6)N$ other throws, so that, on the average, any two aces will be separated by 5 throws, thus making

$$l=5a, \quad \text{and} \quad b=l,$$

as Boltzmann stated; while according to Korteweg, since the value $5a$ for l is obviously correct, we should have found for b the answer $5a/2$. The fallacy in the reasoning of Korteweg, as Clausius pointed out, lies in the fact that if we take points at random on the distance L, there is a much better chance of striking a point on an exceptionally large x than on a small x, because all points of L (not all values of x) are equally probable; the reasoning of Korteweg does not take this into account.

If we now pass from Boltzmann's analogy to the actual case of the gas molecule, we may reason as follows. Divide the long distance L into a very large number Λ of distance elements a. If there are k collisions, we have

$$kl=k\lambda a=\Lambda a,$$

where λ is the number of elements a in l.

If we consider the individual Λ elements in the entire distance, we see that a certain number of them (*viz.*, $k=\Lambda/\lambda$) are terminated by a collision, while in the case of the other $\Lambda-\Lambda/\lambda$ elements the molecule continues at the end of the element without a collision. Hence the probability p of a free path a without a collision at the end of it is

$$p=\frac{\Lambda-\Lambda/\lambda}{\Lambda}=1-\left(\frac{1}{\lambda}\right), \tag{4}$$

while the probability of a collision at the end of any distance element a is

$$\frac{\Lambda/\lambda}{\Lambda}=\frac{1}{\lambda}=1-p.$$

If we now calculate the value of b by the mean value theorem as above, we get

$$b=a(1-p)+2ap(1-p)+3ap^2(1-p)+\cdots$$
$$=[a(1-p)/p](p+2p^2+3p^3+\cdots),$$

or $b=[a(1-p)/p][p/(1-p^2)]$

$$=a/(1-p)=a\lambda=l. \tag{5}$$

If, instead of quantizing the length L into small elements a, we use the continuity method of the calculus, the calculation is as follows. In the expression

$$b=\sum nap^{n-1}(1-p)=\sum nap^n(1-p)/p \tag{6}$$

we write the length x in place of the length na, the factor dx/l in place of

$$(1-p)=1/\lambda=a/a\lambda;$$

similarly for p we put

$$p=1-\left(\frac{1}{\lambda}\right)=1-\left(\frac{a}{a\lambda}\right)=1-\frac{(na/n)}{l}$$
$$=1-(na/nl)=1-(x/ln),$$

and hence for p^n we put

$$\lim_{n\doteq\infty}\left(1-\frac{x}{l}\frac{1}{n}\right)^n=e^{-x/l}.$$

If on substituting we put the factor p in the denominator equal to unity, we get

$$b=\int_0^\infty x(dx/l)e^{-x/l}, \tag{7}$$

which is a well-known expression in the kinetic theory, and is equal to l.

There is another method of obtaining the value of b which is not without interest. If again we divide the entire distance L into Λ elements, each of length a, and take the point to be selected at random successively at the beginning

of each element, then all these points are equally probable, the probability of striking any one of them being $1/\Lambda$. Now, if we selected the first point, the distance traveled to the next collision would be $y=x_1=n_1a$. Hence this particular value of y multiplied by the probability of its occurrence is n_1a/Λ. If we select the second point the corresponding expression is $(n_1-1)a/\Lambda$, and so on, until at the beginning of the last element of x_1 the expression becomes a/Λ. For the next point, the end of x_1, or what is the same thing, the beginning of x_2, we get n_2a/Λ, etc.

Applying now the mean value theorem we find

$$\begin{aligned} b=&n_1a/\Lambda+(n_1-1)a/\Lambda+(n_1-2)a/\Lambda+\cdots+a/\Lambda \\ &+n_2a/\Lambda+(n_2-1)a/\Lambda+\cdot\ \cdot\ \cdot\ \cdot\ \cdot\ \cdot+a/\Lambda \\ &+\cdot\ \cdot\ \cdot\ \cdot\ \cdot\ \cdot\ \cdot\ \cdot\ \cdot\ \cdot\ \cdot\ \cdot\ \cdot\ \cdot\ \cdot \\ &+n_ka/\Lambda+\cdot\ \cdot\ \cdot\ \cdot\ \cdot\ \cdot\ \cdot\ \cdot\ \cdot\ \cdot+a/\Lambda, \end{aligned}$$

and, summing by rows

$$b=\left(\frac{a}{\Lambda}\right)\left[\frac{(n_1+1)}{2}n_1+\frac{(n_2+1)}{2}n_2+\cdots\frac{(n_k+1)}{2}n_k\right],$$

or
$$b=(a/2\Lambda)[\textstyle\sum n_i^2+\sum n_i]=(a/2\Lambda)\sum n_i^2, \tag{8}$$

since the second sum, the value of which is Λ, may be safely neglected; for, depending, as it does, on the first power of the large numbers n it is of a smaller order of magnitude than the first sum, which depends on the squares of the numbers n.

At this stage an interesting fallacy may arise, to which the writer's attention was directed by a question put by a student. If, in order to obtain the customary notation of the theory of fluctuations, we denote the average value of n_i by ν instead of λ, so that

$$k\nu=\textstyle\sum n_i,$$

and if we put for every n_i

$$n_i=(n_i-\nu)+\nu$$

and
$$n_i^2=(n_i-\nu)^2+2(n_i-\nu)\nu+\nu^2,$$

then
$$\textstyle\sum n_i^2=\sum(n_i-\nu)^2+2\nu\sum(n_i-\nu)+\sum\nu^2.$$

Here the last term is surely equal to $k\nu^2$, and the middle sum will be zero, since the algebraic sum of the departures $(n_i-\nu)$ of the n's from their average value ν will cancel. If now we replace the sum of the square terms $(n_i-\nu)^2$ by k times their average value, which we may call β,

$$\textstyle\sum(n_i-\nu)^2=k\beta, \tag{9}$$

and then put β as "mittleres Schwankungsquadrat" equal to ν, our sum is

$$\textstyle\sum n_i^2=k\nu+k\nu^2;$$

and, disregarding the first power of ν in comparison with the second, we should get

$$b=(a/2\Lambda)k\nu^2=a\nu/2=l/2,$$

since $k\nu$ is equal to Λ and $a\nu$ is equal to l. This result is incorrect, however, because the step involving the rule for the mean square of the fluctuations is not permissible. If the foundation of this rule is examined,[6] it will be seen that the quantities n_1, n_2, etc. which measure the lengths of different paths cannot be identified with the n_1, n_2, etc. which measure the probability of the occurrence of an event in a series of N trials.

The proper procedure will now be stated briefly. Returning to Eq. (8),

$$b=(a/2\Lambda)\textstyle\sum n_i^2, \tag{8}$$

we find the $\sum n_i^2$ as k times the average value of the square,

$$\textstyle\sum n_i^2=k(n_i^2)_{\text{Av}}. \tag{9a}$$

and the average value is to be found, as in previous cases, by summing up all possible values, each one multiplied by the probability of its occurrence. We use the same method for obtaining the probability of occurrence of any n^2 (which is the same as that of the corresponding n) as previously, and, taking the probability of occurrence of any n^2 as $p^{n-1}(1-p)$, we get

$$\textstyle\sum n_i^2=k\sum n^2p^{n-1}(1-p)$$

or
$$(\textstyle\sum n_i^2)(p/(1-p)k)=\sum n^2p^n. \tag{10}$$

Now $\sum n^2p^n$, for values of n running up to large numbers, and values of p less than unity, is, with sufficient accuracy[7] given by
$$\textstyle\sum n^2p^n=(p+p^2)/(1-p)^3,$$
and hence
$$\textstyle\sum n_i^2=k(1+p)/(1-p)^2.$$

If we now put

$$1+p=2-(1/\lambda)\quad\text{and}\quad(1-p)=1/\lambda,$$

we get
$$\textstyle\sum n_i^2=k(2\lambda^2-\lambda)\doteq 2k\lambda^2$$

and hence
$$b=a2k\lambda^2/2\Lambda=l. \tag{11}$$

There remains only to show how the last calculation would be made if the calculus were used. Return to the expression

$$b=\frac{a}{2\Lambda}\sum n_i^2=\frac{1}{2L}\sum(an_i)^2=\frac{1}{2L}\sum x^2.$$

Replace $\sum x^2$ by $k(x^2)_{\text{Av}}$. In order to obtain the average value of x^2 by summing up over all possible values, each value to be multiplied by the probability of its occurrence, we use the same procedure as above for $\bar{x}$, so that we get

$$\sum x^2=k\int_0^\infty x^2\left(\frac{dx}{l}\right)e^{-x/l}=2kl^2.$$

Hence we find for b,

$$b=2kl^2/2L=l, \tag{12}$$

as previously.

Accordingly, the two mean paths l, obtained by counting from one collision to the next, and b, obtained by counting from an arbitrary instant to the next collision, are equal. If the two occur together in the discussion of a problem, their values may be interchanged without introducing any error.

[6] See e.g. the summary by R. Fürth, Physik. Zeits. 20, 303 (1919).

[7] G. Chrystal, *Algebra*, Vol. 1 (1886), p. 476.

The Relationship Between Scores on the Scholastic Aptitude Test and College Grades in Physics

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IN 1926 the College Entrance Examination Board devised an aptitude test known as the *Scholastic Aptitude Test* (S.A.T.). It is a test offered to colleges and universities for the purpose of testing all candidates for admission to these institutions. Acceptance, however, is entirely on a voluntary basis, for no institution is obligated to give the test. From 1930 until 1935, the period covered by this investigation, the test consisted of two sections: one in which the material is essentially linguistic or verbal in nature, and another in which the material is essentially mathematical in nature. The test aims to measure potential ability to do work on the college level along specific lines rather than to measure ability in general. A primary purpose is its use as an aid in guidance and teaching in college; that is, in selecting students for admission to the various fields of study, guiding them in their choice of majors, forestalling through supervision the possible failures of doubtful students, diagnosing the mediocre performance or failures of promising students, and selecting groups for instructional purposes.[1]

The S.A.T. was first given to applicants for college entrance by the College Entrance Examination Board in 1926. From 1926 to 1929, inclusive, it consisted only of a verbal, or nonmathematical, section, the composition of which changed during these years. This verbal section at first consisted of nine sub-tests but from 1930 to 1936 consisted of only three sub-tests, totalling 200 items, namely, antonyms, double definitions, and paragraph reading.

Not long after its inception, it became evident that the test was better adapted to selecting students for those subjects in which verbal abilities are required, such as history, English, Latin, than for the mathematical and scientific courses, where mathematical abilities are essential. It was felt that the verbal section alone was not sufficient, because it measured linguistic ability only and not the ability that is thought of as the mathematical and scientific.[2] The dissatisfaction became sufficiently acute to move the Commission on Scholastic Aptitude Tests of the College Board to add a second section, in 1930, to the S.A.T., after experimenting with such a section in 1928 and 1929. This second part consisted of 100 items, arranged in the order of difficulty, dealing with arithmetic, algebra, and geometry, and designed better to select students for the mathematical and scientific courses.

The problem. Since a considerable amount of time has been, and is being, spent on the S.A.T., it seemed of interest to make a study of the degree of relationship between the scores made by students on the mathematical section at the time of their entrance to the University of Pennsylvania and their later academic success, as judged by their marks, in specific courses in elementary college physics. At the same time, the relationship between the scores made by the same students on the verbal section and their marks in the same courses were also determined, thus making it possible, in addition, to compare the relative merits of the two sections as far as elementary physics is concerned. These relationships were investigated for the period 1930–35, during which years both sections of the S.A.T. were given.

The situation is an individual one in each university, so that results obtained at one university are not necessarily directly comparable with similar results obtained elsewhere. Yet, the findings at the University of Pennsylvania, in the sense that they indicate trends, should be of value to other interested groups.

The courses studied. Five one-semester physics courses were included:

P. 1, *Dynamics and Heat*, the first half of a course in general physics.

P. 2, *Sound, Light, Magnetism, and Electricity*, the second half of the general course; prerequisite, P. 1.

[1] E. G. Bill and H. Pennypacker, "Two Deans Consider the Scholastic Aptitude Test," Independent Educ. 3, 17 (1929).

[2] *Annual Report of the Commission on Scholastic Aptitude Tests* (1929), p. 38.

TABLE I. *Distribution of cases by courses in each academic year, 1930–35.*

COURSE	30–31	31–32	32–33	33–34	34–35	TOTAL
P. 1	102	129	131	126	63	551
P. 2	95	115	115	98	56	479
P. 16	84	96	88	92	63	423
P. 17	83	96	88	91	63	421
P. 18	67	74	78	76	51	342
Total	427	510	500	483	296	2216

P. 16, *Elementary Dynamics*, prerequisites, entrance physics, analytic geometry, and enrolment in first calculus.

P. 17, *Properties of Matter, Heat, Sound, and Light*, taken concurrently with P. 16.

P. 18, *Electricity and Magnetism*, prerequisites, P. 16, 17, and enrolment in integral calculus.

The data. All the data for the study were obtained directly from the records of the individual students. Likewise, as far as the University of Pennsylvania is concerned, the data do not represent a sampling; that is, all students were included who had been admitted to the University from September, 1930 to February, 1935, and who, up to June, 1936, had taken one or more of the five physics courses (since the physics courses are sophomore courses, the students who entered during the academic year 1934–35 did not take physics until the following year).

To obtain the data, all the undergraduate schools of the University were visited and the records of their former and present students examined. These included liberal arts, fine arts, education, science, and finance and commerce. When the record of a student was found who had been admitted during the period indicated and who, up to June, 1936, had taken one or more of the five physics courses, his name, school, those of the five courses studied and final mark in each course were noted. The scores in the verbal and mathematical sections of the S.A.T. were obtained from the admission slips. In this way a total of 2216 cases, or marks, were obtained. Of these cases, the percentages occurring in the various colleges and schools were: liberal arts for men, 45; scientific school, 39; electrical engineering, 12; liberal arts for women, 2; education, 2. There were no cases in fine arts and only two in finance and commerce. Except for some students from liberal arts, P. 16, 17, and 18 were comprised almost entirely of students from the scientific and engineering schools.

Statistical treatment. The relationship between test scores and course marks was determined by calculating the coefficients of correlation between the variables involved by the Pearson product-moment method. Ten such correlations were made for each year of admission—five between the verbal test scores and marks in each of the courses and five between the mathematical test scores and marks in each of the same courses. A similar set of ten correlations was calculated for the entire period 1930–35, regardless of the year of admission. Separate correlations by years were made in addition to those for the entire period because it was felt that the latter, while valuable as a summary of what had happened, would not disclose any trends or variations that might have occurred during the years. It will be noticed, likewise, that separate correlations were calculated between test scores and marks in each course, rather than between test scores and the average marks of all the courses; the latter procedure would have yielded a confused picture of the relationships sought.

The number of cases that were available for each of the correlations by years and for the entire period may be determined from Table I. The number of cases for a course in a particular year may be obtained from the column for that year, and for the entire period 1930–35 from the last column on the right.

TABLE II. *Correlations between marks in physics courses and scores on verbal and mathematical sections of S.A.T.*

	1930–31				1931–32				1932–33				1933–34				1934–35			
	V.S.		M.S.		V.S.		M.S.		V.S.		M.S.		V.S.		M.S.		V.S.		M.S.	
COURSE	r	P.E.	r	P.E.	r	P.E.	r	P.E.	r	P.E.	r	P.E.	r	P.E.	r	P.E.	r	P.E.	r	P.E.
P. 1	.379	.058	.367	.058	.226	.056	.376	.051	.353	.052	.513	.043	.435	.049	.547	.042	.322	.075	.483	.064
P. 2	.257	.065	.223	.066	.220	.060	.445	.050	.319	.057	.438	.051	.363	.060	.451	.054	.323	.081	.522	.066
P. 16	.416	.061	.541	.052	.296	.063	.325	.062	.368	.062	.470	.056	.451	.056	.448	.056	.376	.073	.553	.058
P. 17	.439	.060	.481	.056	.330	.062	.264	.064	.279	.065	.485	.055	.589	.047	.420	.058	.397	.071	.555	.058
P. 18	.416	.071	.443	.068	.516	.057	.262	.073	.184	.073	.388	.064	.487	.059	.410	.064	.037	.095	.314	.085

TABLE III. *Means of course marks and of the scores on verbal and mathematical sections for each year.*

Course	1930–31			1931–32			1932–33			1933–34			1934–35		
	X'	V.S. Y'	M.S. Y'	X'	V.S. Y'	M.S. Y'	X'	V.S. Y'	M.S. Y'	X'	V.S. Y'	M.S. Y'	X'	V.S. Y'	M.S. Y'
P. 1	4.3	40.5	40.5	4.7	51.5	44.5	4.4	44.0	38.0	4.6	40.5	38.0	5.1	47.5	49.0
P. 2	4.8	42.0	41.0	5.5	51.5	45.5	5.1	43.5	37.0	5.1	41.5	40.0	5.7	47.5	48.5
P. 16	3.2	49.5	57.5	3.4	46.5	58.0	4.1	46.5	57.5	3.7	46.5	55.5	3.9	47.0	59.5
P. 17	3.9	50.0	58.0	3.8	46.5	58.0	4.5	46.5	57.5	4.4	46.5	56.5	4.5	47.0	59.5
P. 18	4.6	52.0	63.0	4.8	47.0	61.0	4.0	48.0	59.5	3.5	48.0	58.0	4.1	48.0	62.5

TABLE IV. *Standard deviations of course marks and standard deviations of scores on verbal and mathematical sections.*

Course	1930–31			1931–32			1932–33			1933–34			1934–35		
	σ_x	V.S. σ_y	M.S. σ_y	σ_x	V.S. σ_y	M.S. σ_y	σ_x	V.S. σ_y	M.S. σ_y	σ_x	V.S. σ_y	M.S. σ_y	σ_x	V.S. σ_y	M.S. σ_y
P. 1	2.29	28.90	27.50	2.64	25.95	27.90	2.52	27.15	27.90	2.54	28.55	27.10	2.53	28.45	27.40
P. 2	2.25	29.25	27.20	1.97	25.70	27.45	2.01	27.25	26.65	1.91	28.35	27.05	2.04	28.25	27.45
P. 16	2.54	28.70	27.35	2.27	28.15	26.15	2.32	28.85	29.85	2.34	28.75	26.60	2.59	27.55	26.75
P. 17	2.57	28.45	27.00	2.54	28.15	26.15	2.50	28.85	29.85	2.41	28.30	26.55	2.64	27.65	26.75
P. 18	2.41	29.15	24.05	2.52	28.90	25.35	2.48	29.30	29.35	2.65	28.80	26.45	2.43	25.20	25.50

Results and conclusions. The results of the study appear in Tables II–VII. The following abbreviations are used in the tables: P., physics; V.S., verbal section; M.S., mathematical section; *r*, coefficient of correlation; *P.E.*, probable error of coefficient of correlation; X', mean of course marks; Y', mean of scores on verbal section or mathematical section of the S.A.T.; σ_x, standard deviation of course marks; σ_y, standard deviation of scores on verbal or mathematical section of S.A.T.

Analysis by years (Table II):

1. On the whole, the correlations on the verbal section compare rather favorably with those on the mathematical section.

2. For the 1930–31 group, the correlations on the verbal section are about equal to those on the mathematical section in P. 1 and 2. The correlations, however, are low on both sections for the latter course. For the same group in P. 16, 17, and 18 the correlations on the verbal section are somewhat lower than those on the mathematical section, particularly in P. 16, although in all three courses the correlations on the verbal section are very good, above 0.4.

3. For the groups since 1930–31, the correlations on the mathematical section are higher than those on the verbal section in all the physics courses, with the exceptions of P. 17

TABLE V. *Summary of correlations between marks in physics courses and scores on the verbal and mathematical sections for entire period 1930–35.*

Course	Verbal section		Mathematical section	
	r	*P.E.*	*r*	*P.E.*
P. 1	.364	.025	.477	.022
P. 2	.309	.028	.422	.025
P. 16	.374	.028	.441	.026
P. 17	.403	.027	.428	.027
P. 18	.346	.032	.369	.032

TABLE VI. *Summary of means of course marks and means of scores on verbal and mathematical sections for entire period 1930–35.*

Course	X'	V.S., Y'	M.S., Y'
P. 1	4.6	44.5	41.5
P. 2	5.2	45.0	41.5
P. 16	3.7	47.0	57.5
P. 17	4.2	47.0	58.0
P. 18	4.2	48.5	60.5

TABLE VII. *Summary of standard deviations of course marks and of the scores on verbal and mathematical sections for entire period 1930–35.*

Course	σ_x	V.S., σ_y	M.S., σ_y
P. 1	2.42	28.05	27.85
P. 2	2.06	27.95	27.40
P. 16	2.41	28.50	27.40
P. 17	2.54	28.40	27.20
P. 18	2.55	28.55	26.45

and 18 for the 1931–32 group and P. 16, 17, and 18 for the 1933–34 group. All told, the correlations on the mathematical section are higher than those on the verbal section in 18 of the 25 physics correlations.

4. Of the 25 correlations between each section and the physics courses, the verbal section has 2 above 0.5, 6 between 0.4 and 0.5, 10 between 0.3 and 0.4, and 7 below 0.3. The mathematical section, on the other hand, has 6 correlations above 0.5, 11 between 0.4 and 0.5, 5 between 0.3 and 0.4, and only 3 below 0.3.

5. In most instances the correlations for the second semester of a course are lower than those for the first semester. Thus, 3 of the 5 correlations between the mathematical section and P. 2, and 4 of the 5 correlations between the verbal section and P. 2, are lower than the corresponding ones for P. 1. Similarly, all 5 of the correlations between the mathematical section and P. 18, and 2 of the 5 correlations between the verbal section and P. 18, are lower than the corresponding ones for either P. 16 or P. 17, both of these courses being taken concurrently before P. 18.

Analysis for the entire period 1930–35 (Table V):

1. The correlations on the mathematical section are higher than those on the verbal section for every course, when considering results over the entire period.

2. The correlations on the mathematical section are all above 0.4, with the exception of P. 18.

3. The correlations for the second semester of a course are lower than those for the first semester on both sections. This is apparent when the correlations are compared for P. 1 and P. 2, and for P. 16 and P. 17, which are taken concurrently with P. 18. The exceptions to this trend in the correlations for the individual years are few.

4. When it is considered that we are dealing with a single instrument and on the college level, where the student personnel is more homogeneous than on the secondary level, the fact that the correlations for the mathematical section are all above 0.4, with one exception, indicates a considerable positive relation between the scores made by students on the mathematical section of the S.A.T. at the time of entering the University and their future academic success in specific courses in elementary physics. It must also be remembered that the coefficients of correlation obtained are lower than they otherwise would be because not all the candidates who apply for admission and take the examination are admitted, the poorer ones being rejected.[3] The more homogeneous the group, the greater is the restriction in the range of ability, and this restriction tends to lower the coefficients of correlation.[4]

5. Beginning with the June, 1936, S.A.T., taken mainly by candidates for admission in September, 1936, the College Entrance Examination Board dropped the mathematical section. This did not affect the present study which is concerned with students admitted while the verbal and mathematical sections did not change and were taken by all applicants. In place of the mathematical section, the Board is now offering a new series of mathematical examinations, called Alpha, Beta, and Gamma.[5] Since the Board cannot compel the acceptance of its offerings, it remains for the individual university to decide upon their acceptance or rejection.

The results obtained on the mathematical section in this investigation show sufficient promise to indicate the desirability of having such a section in the S.A.T. for the guidance and teaching of students in courses in physics. Since the original mathematical section is no longer available, and since the Board is now offering the new mathematical examinations, Alpha, Beta, and Gamma, in its place, it is suggested that these new examinations be given a trial. It would remain, however, for a future study similar to this one to determine whether the new mathematical examinations are more closely related to marks in physics than the original mathematical section studied here.

6. It will be noticed from Table V that the correlations on the verbal section are not very far behind those for the mathematical section. To determine the reliability of the difference between

[3] *Twenty-Sixth Annual Report of the College Entrance Examination Board* (1926), p. 8.

[4] R. K. Byrns, "Scholastic Aptitude and Freshman Achievement," Sch. and Soc. **35**, 718 (1932); A. B. Crawford, "Forecasting Freshman Achievement," Sch. and Soc. **31**, 129 (1930); F. S. Freeman, "Predicting Academic Survival," J. Ed. Research **21**, 116 (1931); D. Segel, *Prediction of Success in College* (U. S. Office of Educ. Bull. 1934, No. 15), p. 19.

[5] Report of the commission on Examinations in Mathematics, Math. Teacher **28**, 154 (1935).

the obtained coefficients of correlation for the verbal and mathematical sections, the ratio of the difference of the two coefficients to the probable error of their difference[6] was calculated for each course. For P. 1 the chances are[7] 99 in 100 that the obtained difference represents a true difference greater than zero; for P. 2 the chances are 98 in 100; for P. 16, 88 in 100; for P. 17, 67 in 100; and for P. 18, 63 in 100.

In view of the favorable showing of the verbal section in comparison to the mathematical section, it seems advisable, in the case of physics, to use the verbal section in conjunction with a mathematical section.

[6] $\dfrac{D}{P.E.(\text{diff. } r_1 - r_2)}$.

[7] H. E. Garrett, *Statistics in Psychology and Education*, p. 135, Table XV.

Report of the Committee on the Teaching of Physics for Premedical Students*

TO any one familiar with the various published articles and reports on the teaching of physics for premedical students,[1] it is at once apparent that opinion is not divided in any essential or important detail. Medical school authorities and teachers of physics are alike convinced of the great value and importance of physics in the education of the physician. From time to time, however, controversial opinions have been expressed concerning minor points. The committee lists some of these points in the following paragraphs, together with its opinion or opinions.

1. *Should premedical students be compelled to pass the same "First Course in Physics" that introduces the pre-engineering students to that subject? Or should a special (presumably easier, or at least less mathematical, course be presented, in which particular attention would be paid to biologic and medical applications?*

The first course in physics for the premedical student is of tremendous importance. It should be carefully designed to give the student as thorough a grasp as possible of general physics. The introduction of material especially selected for its medical or biological content must be considered as of minor importance. In those colleges that require an unusual mathematical background (calculus, for example) as a prerequisite to the first physics course for engineering students, it may be necessary to set up a course for the premedical student in which the approach is different. However, such a necessity must not be permitted to reduce the content of the premedical course in any essential way.

2. *Is the present one-year course in general physics sufficiently comprehensive for the prospective student of medicine?*

The committee believes that the time has come to face the facts with regard to this important question. No student, except he be a physics major, needs physics to a greater degree than the prospective physician, for no other subject can compete with physics in the development of those habits of exact thinking and inquiring self-criticism which are so important to the modern physician. Furthermore, physics is one of the required subjects which are necessarily completed during the premedical years; it is not taught, or even reviewed, in medical school. Some of the subjects that are now taught during the premedical years are, on the other hand, so thoroughly covered during the medical school years as to make their somewhat haphazard presentation to the premedical group appear superfluous, to say the least. We refer to such subjects as biology, comparative anatomy and embryology. We are informed that the teachers of anatomy, histology, and physiology in medical schools are mostly of the opinion that their work is not made appreciably easier for them because of the student's premedical studies in these medical school subjects.

Under date of June 11, 1936, a high authority in the field of medical education wrote to a member of this committee as follows (we quote the last two paragraphs of the letter):

> "There has been much complaint on the part of medical colleges that students do not know how to use apparatus or judge results from its use because they have not had the right sort of course in physics. Has the physics teacher set up a special course for premedical students? Or has he not had sufficient time to give a good course in physics? If the former, he has made a serious mistake. If the latter, then why not insist that the course should be longer than one year? The teacher of physics must determine what shall be done. We insist on *not less than* one year. That's all we prescribe.
>
> "I trust that your association will make this a topic for discussion at its next annual meeting and if it does, I would like to know the result of the discussion. A paper for publication in our *Journal* will be most acceptable."

The committee feels that this is a challenge and that the American Association of Physics Teachers should go on record as in favor of making the physics prerequisite two years instead of one: the first year's course to be either

* Reprints of this report may be obtained from the Editor, the cost of 6 reprints being 30 cts. postpaid. Reprints of *Physics in Relation to Medicine* (1923) are also available at 10 cts. per copy.

[1] "Physics in Relation to Medicine," a Report by the Educational Committee of the American Physical Society (1923); reprinted in Am. Phys. Teacher 2, 48, 101 (1934).

E. L. Harrington, "On the Question of Physics for Students of Biology and Medicine," J. Assoc. Am. Med. Coll. 7, 362 (1932); "On Physics in Relation to Medicine," Am. Phys. Teacher 2, 176 (1934).

G. W. Stewart, "Heresy Concerning Specialized Physics Courses," Am. Phys. Teacher 1, 65 (1933).

"Report of the Committee on Differentiation in First Year Courses" (particularly paragraph 6), Am. Phys. Teacher 2, 33 (1934).

A. E. Caswell, "The Content of the First Year Course in College Physics," Am. Phys. Teacher 2, 95 (1934).

K. K. Smith, "Physics and the Premedical Student" (part of a Symposium on Instruction for Premedical Students), Assoc. Am. Med. Coll. 11, 145, (1936).

R. K. Cannan, "The Physical Sciences in the Training of the Physician" (part of a Symposium on Instruction for Premedical Students), J. Assoc. Am. Med. Coll. 11, 173 (1936).

W. E. Chamberlain, "The Extraordinary Importance of Physics in the Education of the Physician," submitted for publication in J. Assoc. Am. Med. Coll.

identical with, or of at least equivalent content to, the regular first course for engineering students and physics majors; the second year's course to be specially designed for the embryo physician, not with the aim of making it easy for him to get over the additional hurdle, but for the purpose of assuring the hard-pressed medical school faculties of a proper supply of well-educated and educable medical students.

In this connection it should be noted that less than 60 percent of those applying to the medical schools for the first year standing are accepted; in 1935 only 6150 actually matriculated and 45.9 percent of the 12,740 applicants were definitely rejected by the medical schools. The teachers in the premedical years are but doing their part in an important work if careful grading, in advanced as well as in elementary courses, renders the selection of the first year class in medicine more decisive and accurate. This is especially true in the case of physics, for that subject has been specified by medical school faculties and by the Association of American Medical Colleges, as of such importance that a student cannot enter a medical school in the United States or Canada unless he has a satisfactory grade in college physics.

3. *What of physics electives for premedical students?*

It is obviously our duty to offer elective courses, especially while the required course is inadequate, as at present. That a majority of students have time available for such electives is apparent from figures presented in the *Journal of the Association of American Medical Colleges*, **11**, 185 (1936), by Dr. Fred C. Zapffe, Secretary of that Association. Eighty percent of the accepted applicants for first year standing in the medical schools in 1935 had three years or more than three years of premedical college work. Yet many hundreds of successful applicants fulfilled the minimum requirements in but two years. Obviously, after minimum requirements have been met, the three and four year students have time available for physics electives.

4. *Should physics be given a place in the curriculum of the medical school?*

The committee feels that it would be presuming for recommendations along these lines to emanate from the physicists. At the same time no one has refuted, or even objected to, the statements made in 1923 in the report, "Physics in Relation to Medicine," of the Educational Committee, American Physical Society.[1] Presumably it is still true that there is a strong demand "not only for increased preparation in mathematics and physics for the student entering the medical school, but for increased participation of physicists in medical research." The need for attaching physicists to the faculties of medical institutions is at least as great as it was in 1923, though the number of physicists thus engaged is still small. The 1923 report was probably wise in its recommendation that "there should be provided, in the medical school, instruction in physics as directly applied to medicine, as part of the duties of the physicist recommended in another section of this report."

In conclusion, This committee has been in existence for three years and we are by no means certain that our continuation is essential. It would seem that a new personnel might have advantages, or at least a fresh point of view. Furthermore, we have recently learned that authorities in the field of medical education tend to object to the term "premedical" as applied to students preparing to enter medical school; their claim is that "regimentation" of such students under the special category "premedical" tends to narrow their opportunities, substituting narrow "training" for broad "education." This suggests a change of name, or an entirely new committee. With these thoughts in mind the committee makes the following recommendations:

1. That the present "Committee on the Teaching of Physics for Premedical Students" be discontinued.
2. That careful consideration be given to the appointment of a new committee, possibly to be known as the "Committee on Physics in Relation to Medical Education."
3. That if such a committee be appointed, it should have as its objectives: (a) to continue, in collaboration with the editor of the *Manual of Demonstration Experiments*, the work of collecting, selecting, and developing experiments which will illustrate physical principles but at the same time have sufficient physiological application to make a direct appeal to the student of medicine; (b) to continue, by correspondence with medical school authorities and cooperation with the Association of American Medical Colleges and the Council on Medical Education and Hospitals of the American Medical Association, to study the special needs of medical institutions and students of medicine in the field of physics and its teaching.

W. EDWARD CHAMBERLAIN, *Chairman*

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Questionary on the Metric System

PHYSICISTS who are interested in promoting and guiding the metric movement in the United States may obtain a printed questionary from the Metric Association, Pottsville, Pa., which they are urged to fill out and return to the Metric Association for tabulation.

APPARATUS AND DEMONSTRATIONS

An Automatic Polariscope

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A POPULAR demonstration in physical optics is the one relating to the appearance of various transparent substances when placed between two plane polarizing devices, "crossed" with respect to each other. It is a great convenience for the lecturer to be able to demonstrate this phenomenon before classes without being hampered by the manual details of inserting and removing a number of specimens between polarizer and analyzer. To attain this end, the device shown in Figs. 1 and 2 has been constructed.

A circular disk of aluminum Z, about 24 in. in diameter and $\frac{1}{8}$ in. thick, is mounted to rotate on a horizontal axis. The specimens to be viewed in polarized light are mounted over circular openings cut out near the rim of the disk, and spaced 45° apart. The following specimens have been used: A, a piece of plate glass showing no strains; B, a piece of glass under stress impressed by a screw; C, a U-shaped piece of celluloid under strain to demonstrate the photoelastic art; D, a Cellophane cigar wrapper showing striking color effects; E, a roll of mica sheet; F, thin gypsum plates to form a colored cubic pattern; G and H, specimens of strained glass showing symmetrical stresses.

On opposite sides of the disk at PP are mounted pairs of Polaroid disks of about 4 cm diameter, one in front of and the other behind the disk. Ordinary light bulbs placed in suitable housings SS equipped with lightly frosted windows, serve as light sources. Two other housings LL contain the light sources which enable one to view the specimens in ordinary light. It is very desirable to show the marked contrast in the appearance of a specimen when viewed successively in ordinary and in plane polarized light. The disk Z is rotated by a 1/20-hp motor M. By the use of suitable reducing gears, belt, and wheel W, the time for one uninterrupted rotation of the disk is made about 9 sec.

In order to afford the observer sufficient time to view each specimen as it passes between polarizer and analyzer, an intermittent friction drive is employed. This consists of two wooden

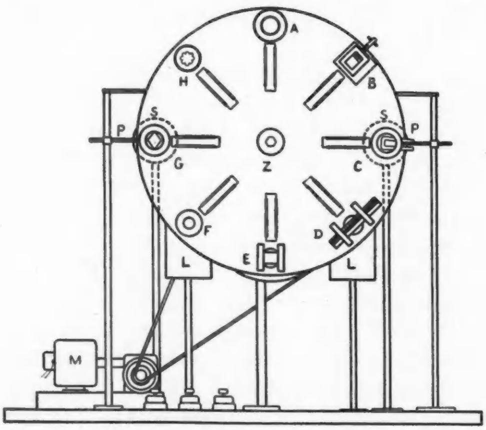

Fig. 1. Front view of polariscope.

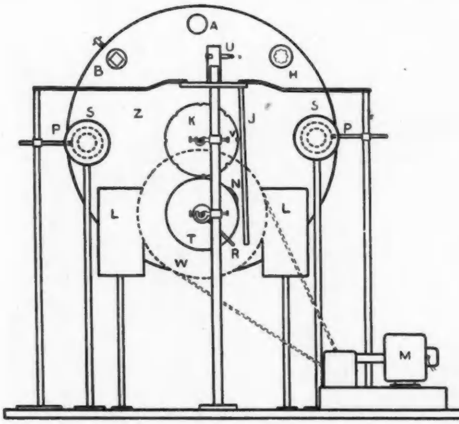

Fig. 2. Rear view of polariscope.

wheels K and T of equal diameters, to the edge of the latter being attached a strip of emery cloth N of sufficient length to turn the disk Z through about 45°. Rotation of the disk each time through exactly 45° is effected by the cam action of the rod R which revolves with the wheel T. Just before the emery strip N engages the wheel K, the rod pushes out the small stop V attached to the spring J, and then allows it to drop back and engage the succeeding small semi-circular groove. These grooves are located at 45° intervals on the edge of the thin metal disk attached to the wooden wheel K. This arrangement assures that the disk Z holding the specimens will always rotate through exactly 45°, and then remain in position for about 8 sec. The complete cycle of events requires about $1\frac{1}{4}$ min. For smooth operation the disk is properly balanced. The jar occurring at the instant the disk stops is materially reduced by employing the friction rubber U (Fig. 2).

The apparatus as described is especially suitable for demonstration before a small class, and as a physics museum piece for the interest of the general public. Labels attached to the various parts assist in the explanation of the phenomena to be observed. However the apparatus can be just as successfully used for demonstrations before large lecture classes. One need but substitute for one of the lamp housings S, an arc lamp and suitable focusing lens for projecting enlarged images of the patterns upon a viewing screen. Obviously the apparatus may be enlarged to house a larger number of specimens if so desired.

Inexpensive Apparatus for Study of the Raman Effect

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APPARATUS commonly used in the study of the Raman effect is so expensive that it is beyond the reach of the average laboratory. In beginning work on this subject, the authors were faced with the necessity of improvising equipment with which results of reasonable accuracy could be obtained. A description of the resulting apparatus may be of interest to those experiencing a similar lack of equipment. Furthermore, the apparatus is composed of parts so universally available as to make possible a more general introduction of experiments on the Raman effect into advanced courses in physics and chemistry.

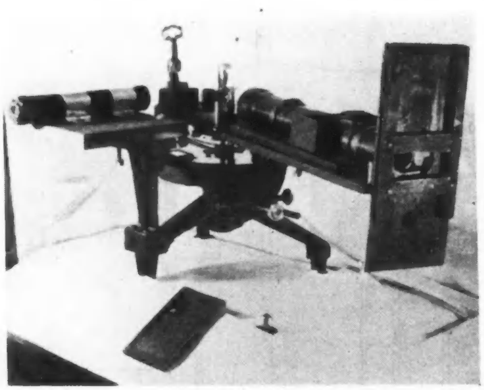

FIG. 1. The spectrograph, showing camera attachment.

Spectrograph. A student spectrometer was changed into the two-prism spectrograph shown in Fig. 1. The telescope of the spectrometer was replaced by a camera attachment consisting of a wooden platform upon which was mounted the tube and objective (aperture 30 mm, focal length 30 cm) of a telescope from a surveyor's transit. The plateholder, which accomodates a 4.5×6-cm plate, is shown in the lower part of the figure.

The two prisms (base 45 mm, height 22.5 mm, polished on two faces) are mounted on a wooden platform rigidly attached to the original prism table. The collimator (achromatic lens, aperture 25 mm, focal length 15 cm) is displaced, parallel

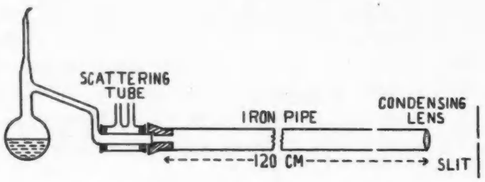

FIG. 2. Diagram of the scattering tube.

to its initial position, and clamped on a wooden platform attached to the original collimator arm in such a manner as to illuminate uniformly the face of the first prism. Both camera and collimator platforms can be leveled by use of the original leveling screws. The dispersion is 40A/mm at 4500A and the length of the spectrum from 4000 to 5000A is 20 mm. The luminosity is sufficient for photographing the six strongest Raman lines of benzene (excited by Hg 4358) in 10 min. when no filter is used.

The spectrograph is mounted on a board containing three leveling screws and is covered with a large, light-tight, cardboard box which is painted a dull black. This box telescopes into a frame on the laboratory table and has a 2 cm hole for admitting light to the slit and a large removable shutter for access to the camera back. This arrangement effectively prevents stray light from reaching the plate.

Scattering tube and accessories. Light from the scattering tube (Fig. 2) is condensed on the slit by the achromatic objective (aperture 25 mm, focal length 15 cm) of the original spectrometer telescope. The scattering tube is of the type designed by R. W. Wood. It is made from Pyrex tubing, 6 mm inside diameter, to the end of which is sealed a disk of glass taken from the bottom of a Pyrex beaker. The other end is bent sharply upward and sealed to a 25 ml distilling flask, as shown in Fig. 2. The length of the straight portion of the Raman tube is 10 cm, thus making it possible to observe the spectrum of a substance when only 3 ml of material is available. By decreasing the length of the tube and increasing the exposure time, one could use still smaller samples. The scattering tube is surrounded by a jacket through which solutions of sodium nitrite or cobalt sulfocyanate may be circulated for the purpose of isolating the desired exciting line and for cooling.

In adjusting the apparatus, one places the iron pipe (Fig. 2) in line with the collimating tube and the telescope tube containing the condensing lens by sighting along a stretched string. A one-hole stopper is then placed in the end of the iron pipe and illuminated; when proper alignment is secured the spot of light will focus accurately on the center of the slit. The one-hole stopper in the end of the iron pipe serves as a diaphragm, and as a support and centering device for the plane end of the scattering tube, which is placed in line by sighting along the iron pipe. A fixed clamp serves to hold the tube in this position and facilitates realignment after distillation.

Mercury vapor lamp. The source of light is a Pyrex mercury vapor lamp (Fig. 3) which was made from tubing having an internal diameter of 1 cm. Since evaporation is more rapid at the positive side, the size of the bulb condenser should be such that the condensation just balances the evaporation; after much experimentation, it was found that the bulb from a 50 ml distilling flask was of the proper size and shape to accomplish this. As a further control, the outlet tube used in the evacuation of the lamp was sealed in such a position as to be just below the normal mercury level; if evaporation becomes too rapid on this side, causing the mercury to drop below this tube, then it serves as a condenser and restores the mercury to its normal level. This arrangement permits continuous operation during long exposures with a minimum of attention. While a Pyrex stopcock is convenient, it is possible to seal any stopcock to the outlet tube with *para* rubber tape at a distance of approximately 5 cm from the junction with the larger tube. Reevacuation of the lamp is desirable whenever it starts with difficulty; this, however, occurs only at infrequent intervals of perhaps two or three weeks.

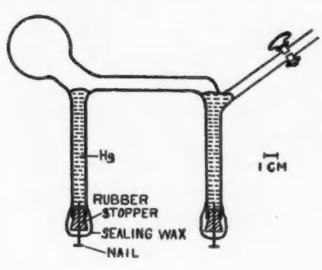

FIG. 3. The mercury vapor lamp.

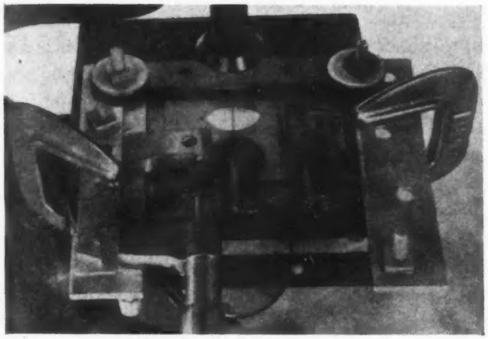

FIG. 4. The improvised comparator; the micrometer caliper pushes the movable glass plate, on which the spectrogram is clamped, between the parallel metal strips.

The lamp as described is good for approximately 100 hr. of operation, after which time the luminous portion of the tube darkens, acquires a different expansion coefficient than that of the nonilluminated Pyrex, and cracks off. (After this change, it is more difficult to fuse the portion which has been illuminated.) It is then necessary to replace the horizontal portion, but this is easily done without disturbing the remainder of the structure. The lamp operates at 3 amp. from a 110-v, d.c. generator with a series resistance of 15 ohms. If lesser voltages are used, the length of the horizontal portion must be correspondingly reduced. Cooling is effected by a blast of air from an electric fan, directed against the bulb condenser and along the axis of the illuminated tube.

The lamp is supported by clamps on an iron stand and is started by tilting. An elliptical mirror made from polished aluminum, supplied by the Aluminum Company of America, focuses the light of the arc at the center of the scattering tube, thus increasing the intensity of illumination many-fold.

Comparator. The plates are measured by the use of a comparator improvised from an ordinary microscope (magnification 10 diameters) and a micrometer caliper, as shown in Fig. 4. A 1.5–2X objective is used in conjunction with a 5X eyepiece which contains a cross hair prepared by the method suggested by A. W. Footer.[1] The caliper is bolted to a metal frame which is in turn clamped to the stage of the microscope. Between the parallel sides of the frame, a 7 cm square piece of glass is pushed by pressure of the micrometer against a piece of metal, glued to the glass. The 4.5×6 cm photographic plate is clamped to this movable glass plate by two brass connectors (Cenco 83867) which have been sawed in two and fastened to the glass plate with Duco cement.

Experimental procedure. In practice, an iron arc comparison spectrum is placed on the plate by removing the condensing lens and shielding that part of the plate where the Raman spectrum appears by means of a wire, 0.8 mm in diameter, placed directly in front of the plate. One end of the strip of wood which supports this wire is visible in Fig. 1. Removing the condensing lens is equivalent to diaphragming the collimating lens, since the source is a distant one, and gives exceedingly sharp iron lines, thus increasing the accuracy of measurement. The iron arc is placed at the end of the iron pipe (Fig. 2) and an exposure time of 1 min. is usually sufficient.

The wire is then removed, the condensing lens replaced, a light-tight connection made with black paper between the condensing lens and the box covering the spectrograph, the scattering tube lined up, and an exposure of from 1 to 20 hr. made. Redistillation at intervals during this time is necessary for some substances. Eastman 40 plates, backed to prevent halation, are used and development is in Eastman D19 solution for 6 min. at 18°C.

The wave-lengths are obtained by linear interpolation from adjacent lines in the iron arc spectrum. The accuracy of measurement as determined by the average deviations from the mean values given by Kohlrausch in *Der Smekal-Raman-Effekt* for benzene, carbon tetrachloride, and phenylacetylene is 3 cm^{-1}. This accuracy is rather surprising when one considers the improvised nature of the apparatus and the inexpensive lenses and prisms employed.

With this apparatus, the Raman spectra of *di-n*-butyl ether and ethyl adipate have been measured.[2]

The work described was supported in part by a grant from the Virginia Academy of Science and the authors wish to express their appreciation of this aid.

[1] Science **84**, 490 (1936).

[2] J. Chem. Phys. **5**, 752(L) (1937).

A Student Type Portable Geiger-Müller Counter

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A SIMPLE form of Geiger-Müller counter is useful in the general laboratory for demonstrating cosmic rays and the nature of radioactivity. As a part of the laboratory work for premedical students, it may be used for elementary experiments in radioactivity.

The G-M counter described is simple, compact and portable; it weighs 52 lb. Its portability and independence of power mains make it useful indoors or out. The counter circuits are enclosed in a single wooden box (Fig. 1), 13×16×11 in., with front, back and instrument panels of Masonite. A part of the front is hinged. When down, it exposes the recessed instrument panel; when up and latched, it protects the panel in transit. The two carefully insulated circuits of the counter are a counting tube circuit, which includes a high voltage battery, and a single-stage amplifier circuit. These circuits are mounted on a shelf behind the instrument panel and above the battery compartment. Both the front and the back of the box are removable so that repairs can be made, and tube or batteries replaced. The back has a heavy glass window through which the counting tube, amplifier and part of the batteries are visible. This makes it easy to examine the fundamental elements of the circuit without danger of exposure to high voltage, and to fix the position of the counting tube relative to the active source.

Twenty compact Burgess type Z30P batteries are used for the high potential source. The 900-v battery weighs 27 lb. and occupies 0.3 ft^3. The counting tube circuit requires less than 1 μamp. so that the battery life is shelf life—ten months for most, better than a year for some. The sensitivity and stability of the counter are due to the excellence of the counting tube C. This tube, designed by G. L. Locher, is of the γ- and cosmic-ray type; it operates on approximately 800 v and can be purchased[1] or easily made in an ordinary laboratory.

The circuit (Fig. 2) is similar to that described by Johnson and Stevenson[2] except that the resistance across the neon glow lamp N, a standard 2-w type, is replaced by an impedance L (the secondary of a small audiotransformer). Impedance causes the lamp to flash on less plate voltage and with brighter flashes. Four of the Z30P batteries connected across a selector switch S, together with a potentiometer V across a medium size B battery, give both rough and vernier control of potential up to a maximum of 945 v with new batteries. The operating potential is not critical, but the vernier control can be used to determine threshold potential and counting range. The potential reserve is used to compensate for loss in service and to demonstrate paralysis of the counter. The coupling unit into the type-30 amplifier tube is an ordinary 10-megohm grid leak. Batteries for the amplifier are ordinary dry cells and medium weight B batteries. The milliammeter $M.A.$, of range 0 to 5 ma, checks operating conditions and dips for each ionizing pulse. The mean level of its readings changes with the activity, decreasing

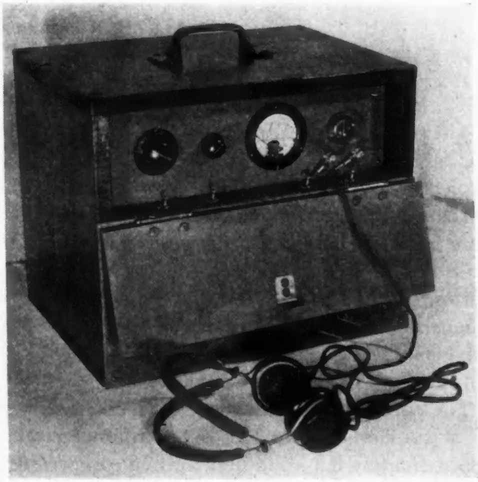

Fig. 1. Photograph of the portable counter.

[1] Herbach and Rademan, Inc., Philadelphia, Pa.

[2] J. Frank. Inst. **216**, 329 (1933).

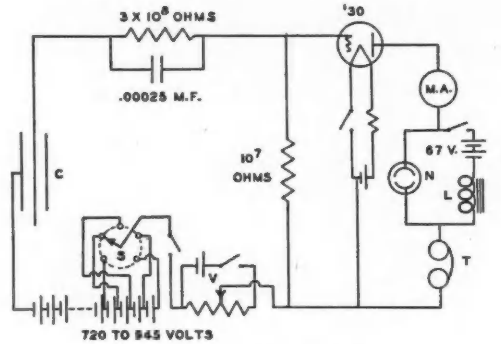

FIG. 2. Counter circuit diagram.

as a radioactive source is brought nearer. When very close, the neon lamp seems to glow continuously and the meter remains fairly constant at a low level. The telephone *T*, connected to binding posts on the control panel, clicks loudly enough to be distinctly heard when placed on the cover of the box.

The count is obtained by watching the flashes of the neon lamp and listening to the clicks of the telephone. With a little practice, using a hand tally counter and stop watch, an observer can record every tenth count over a period of several minutes and compute the counting rate for counts not exceeding 300 per min. The count is only fairly accurate, but the observations bring out facts about the nature of radioactive disintegrations and give some idea of what probability distribution means. Students who wish to extend the work into a project and do more exact counting, may connect a suitable register circuit and message register in place of the telephone.

This G-M unit has been used for the following elementary experiments: (*a*) study of counter tube and associated circuits; (*b*) determination of threshold voltage and counting range; (*c*) study of the activity of the counter as the distance to the active source is changed; (*d*) comparison of strong and weak sources at the same distance from the counter; (*e*) study of the shielding effects of various thicknesses of lead; (*f*) study of methods of searching for lost radioactive materials; (*g*) counting of cosmic rays outdoors; (*h*) comparison of sensitivities of the counter and a γ-ray electroscope.

A Horizontal Projection Cloud Chamber

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IN the classroom or lecture hall a direct projection cloud chamber that can be used with the standard lantern and without a vertical projector has been found very useful.[1] The horizontal projection cloud chamber described here is simple and compact, and consists essentially of a sylphon closed by two glass end plates. At the same time it has the advantage of a consistent, adjustable, mechanically controlled expansion, and is designed to approach the optimum conditions for good tracks.

The chamber consists of two live corrugations of a sylphon 3.5 in. in outer diameter and approximately 0.5 in. in depth when compressed, as shown in Fig. 1. To the sylphon are soldered two brass rings *f*. A glass plate 3 in. in diameter fits into each brass ring and another brass ring *e* can be tightened down on the plate by means of eight screws. The chamber is made tight by two pairs of rubber gaskets *d* placed between the glass and metal plates. The brass ring *f* is machined so that the glass plate sits as far into it as is practicable in order that the chamber may be as shallow as possible. The sylphon must have enough depth to permit the desired expansion and to retain its spring. On the other hand, a large chamber-depth leads to an increase in the necessary expansion distance, poor contrast between tracks and background, and greater turbulence.

The radioactive source can be introduced into the chamber in several ways, one of which is to insert it into one of the brass rings. The device

[1] M. S. Livingston, Am. Phys. Teacher **4**, 33 (1936), described a vertical projection cloud chamber.

shown in Fig. 1 for inserting it through the glass plate also provides a convenient means of introducing the ion-sweep potential, since it serves as

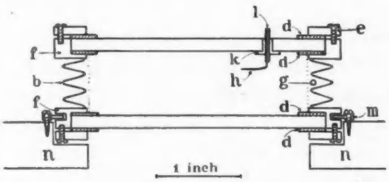

FIG. 1. Cross section of expansion chamber.

one electrode. The upper glass plate has a hole drilled near the circumference, and into this is cemented a brass collar *k* which has a bore threaded to hold a small machine screw. Litharge and glycerin was used as an alcohol and water resisting cement. Apiezon "*Q*" placed in the thread of the screw is successful in keeping the chamber tight. The screw *l* holds the nickel or silver wire *h* on whose tip is the radioactive source, a deposit of RaF (polonium). The wire is bent toward the center of the chamber in order to be visible on the screen. The brass collar *k*, screw *l*, and wire *h* act as one electrode, and the sylphon as the second electrode of the ion collecting potential field. A collecting potential of 120 v is sufficient, but 240 v is desirable. The conventional method of using an Aquadag ring on the upper glass plate as one of the collecting electrodes is of course also successful.

In order to avoid the presence of excess liquid which, especially in the case of alcohol, tends to cause a general fog, a wick *g* consisting of some absorbent cotton rolled into a rope is used.

The turbulence of the expansion is influenced by several factors and may be reduced in many ways. A baffle consisting of a cylinder of 80-mesh wire cloth separating the chamber from the region next to the corrugations was found to decrease the turbulence appreciably. It is kept in position by several of its own thin wires held between the rubber gasket and the brass plate *f*. During the compression it deforms slightly but returns to its original position after the expansion. Turbulence resulting from compressional waves set up by the mechanical shock at the end of the expansion may be reduced by damping the system with a felt or rubber stop under *r*.

The mounting of the chamber is shown in Fig. 2. The chamber *a* is set into a heavy wooden baseboard *n*. Wood screws going through extensions of the brass ring *m*, Fig. 1, hold the chamber firmly on the base. A Duraluminum arm *t*, 8 in. long, bent to form a U, has one end fixed and the other end lowered and lifted. The arm is connected by two pins to the upper ring *f* of the chamber, 4.4 in. from the fixed end. Although expansion may be accomplished by releasing the arm by hand or by a trigger, a cam arrangement is simpler and more satisfactory to use. The cam *s*, mounted with its center $2\frac{1}{4}$ in. above the base, has a drop of $\frac{1}{2}$ in.; this allows for more than the necessary expansion ratio which in this case is of the order of 1.11. A spring *p* is mounted between the base and a cross bar *o*. Although compression of the gas in the chamber is usually sufficient to carry through the expansion without its use, the spring permits control of the rate of expansion and ensures sufficient expansion. A finely threaded rod *q*, coaxial with the spring, goes through a hole in the crossbar. A heavy nut *t* faced with felt or sponge rubber acts as an adjustable and shock-absorbent stop to the expansion.

The cloud chamber can be mounted immediately in front of the usual lantern slide holder or can replace the slide holder by a simple clamp

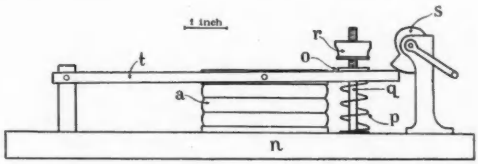

FIG. 2. Lateral view of cloud chamber.

arrangement. No water cell is necessary as a small change in the expansion ratio will compensate for changes in temperature. When the tip of the wire carrying the source is focused on the screen the tracks appear as long black lines against a white background which can be easily seen in a large auditorium. A fairly strong source of RaF is used so that a considerable number of tracks always lie within the region of the focal plane of the optical system.

The suggestions of Dean G. B. Pegram have been very valuable.

Phase Control of Thyratrons—An Experiment for the Undergraduate Laboratory

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THE increasing importance of the thyratron in experimental and industrial applications makes it a fitting subject for a laboratory experiment in an advanced undergraduate course in electronics. Such procedures as obtaining the breakdown-voltage curve and constructing a relaxation oscillator are well known. Although phase control of gaseous triodes is of greater industrial importance, published circuits call for complications and expense outside the possibilities of many courses and institutions.

Within the last few years, the cost of thyratrons has decreased considerably.[1] The smallest thyratron, the RCA type 885, sells for only two dollars and, in spite of its low power rating, it has proven satisfactory for experiments and demonstrations.

There are two common methods which can be employed to demonstrate grid phase shift.[2] One uses a condenser and resistance combination; the other uses a smoothing choke in place of the condenser. The first method is suited for use with the larger tubes, or in applications where the grid must be moved into phase automatically in a predetermined time. Where large loads are to be controlled, or where the load is not constant, it is customary to use a saturable reactor whose a.c. impedance in the load circuit is a function of the thyratron-controlled unidirectional current applied through a second winding.

Figure 1 shows the simplest workable circuit for controlling the intensity of a small lamp by shifting the phase of the grid of a thyratron, using an inductance-resistance circuit. The parts used are of the type which are found in the smallest undergraduate laboratory, but even if all parts, including the RCA type 885 thyratron, have to be purchased, the cost will be less than five dollars. A 5- and a 6.3-v winding may be connected together to obtain the 11.3-v r.m.s. suggested in the diagram. Care must be taken that these windings are connected in series-aiding and in proper phase with respect to the plate voltage. Fig. 2 shows the entire assembly in operation. The phase shift may be observed and measured on a cathode-ray oscillograph[2, 3] if one is available, and the average voltage across the lamp compared with the value calculated from this observed phase shift and the characteristic curve of the tube. Only approximate

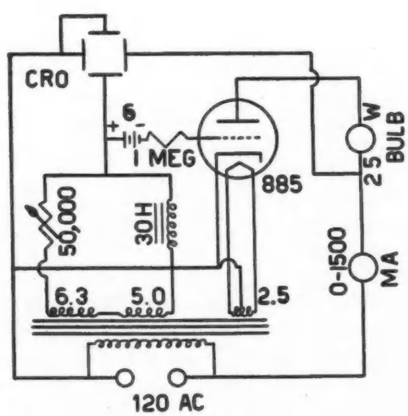

Fig. 1. Schematic diagram of circuit, including cathode-ray oscillograph, to observe phase control of thyratrons.

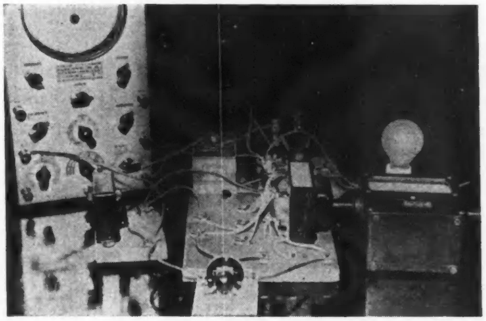

Fig. 2. Photograph of demonstration assembly.

[1] The "seconds" that can be obtained at a special educational price from the General Electric Co. are frequently as useful as the commercially marketed tubes.

[2] Wood, "Phase Measurements with the Cathode-Ray Oscillograph," R. S. I. 2, 644 (1931).

[3] Hughes, "Thyratron Selector for Double Trace Cathode-Ray Oscillograph," R. S. I. 7, 89 (1936).

agreement can be expected in a simple calculation which neglects grid current.

It will be noted from Fig. 2 that we have found it expedient to standardize in the laboratory on individually mounted components with Fahnestock clip terminals.

Stroboscopic Aids in the Teaching of Physics

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Although many stroboscopic experiments are merely interesting, some may also be useful in making quantitative studies of certain phenomena. Recent advances in the design and construction of stroboscopes have increased their flexibility and reliability, and have made it convenient to extend their use in lectures and laboratory experiments. Some such extension has been made and there will be discussed briefly the particular type of stroboscope used, some of the experiments performed, and an example of the results obtained.

Mechanical[1] and electrical[2] devices have been used for producing intermittent flashes of light and either method could be used in many cases, but the electrical method has some advantages over the mechanical method. A simple electrical circuit described by Germeshausen and Edgerton[2] has been used in this work. One of its unique characteristics is the use of a "Strobotron" tube, which has a cold cathode so that no "warm-up" period is required. The duration of the light pulse from the strobotron is of the order of 10^{-5} sec. or less, making it possible to obtain exceedingly sharp stroboscopic patterns. The relaxation oscillator used to excite the strobotron is very simple and has a linear adjustment of frequency, in addition to two fine adjustments, one effective at high frequencies, the other, at low frequencies. The easily portable unit may be made completely a.c. operated and the frequency of light flashes is practically independent of ordinary fluctuations of line voltage.

Four alterations in the aforementioned circuit have been made to adapt it to the present work: (1) two identical stroboscope units operate from the same pulse, thus increasing the light intensity; (2) the relaxation oscillator and the stroboscope are separate units so that the oscillator may be used for other work, or the stroboscope may also be excited by means of, say, a metronome; (3) the low frequency limit is depressed to about 30 pulses/min. with little or no sacrifice in constancy of frequency; (4) a type 5M Western Electric message register is used in the stroboscope in place of the 3500-ohm resistor. The last is the important alteration since it enables one to determine, *very directly*, pulse frequency as high as 1200 per min. This frequency is sufficiently high to include many experiments, but if higher frequencies are desired, calibration is not difficult, or a direct reading frequency meter[3,4] may be used.

All parts of the relaxation oscillator (Fig. 1) may be ordinary radio parts, except that the 10^5-ohm potentiometer should be wire-wound for good control. The condensers C_1 have values 0.01, 0.02, 0.05, and 0.15 μf, so connected that any combination of them may be connected in

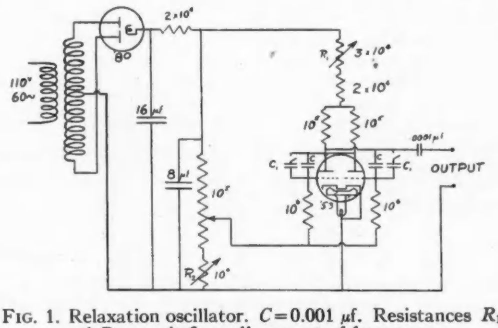

Fig. 1. Relaxation oscillator. $C=0.001\ \mu$f. Resistances R_1 and R_2 permit fine adjustment of frequency.

[1] Bales and Blackburn, Am. Phys. Teacher **5**, 139 (1937).
[2] Germeshausen and Edgerton, Electronics **10**, 12 (1937).
[3] F. V. Hunt, R. S. I. **6**, 43 (1935).
[4] Gingrich, Evans and Edgerton, R. S. I. **7**, 450 (1936).

parallel with C. The stroboscope is shown in Fig. 2. The input condenser may be omitted when the stroboscope is excited by means of the oscillator.

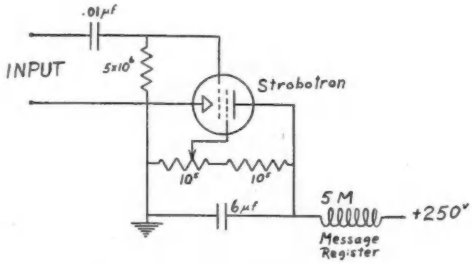

FIG. 2. The stroboscope circuit.

An oscillatory system to be studied with the stroboscope should be one in which the oscillations are sustained or at least one in which the damping is a small factor. The device shown in Fig. 3 provides sustained oscillations when a blast of air from a small fan is directed against the flat side of the oscillating body S. The springs a support S, and S will oscillate with a frequency which depends on its mass and on the stiffness coefficient of the springs. While S is oscillating, stroboscopic light is directed upon it, and coincidence of the frequencies of S and of the stroboscope is easily obtained in a room only slightly darkened. When coincidence is obtained, the counter mechanism of the message register is released for, say, 1 min. The total electrical

TABLE I. *Typical data.*

	S (ALONE)	S WITH 1 ROD	S WITH 2 RODS
Mass (g)	125.5	189.2	252.9
Obs. pulses/min.	809; 810; 808	666; 669; 666	583; 583; 583
Obs. frequency (sec.$^{-1}$)	13.48	11.12	9.72
Calc. frequency		11.17	9.75

pulses in 1 min. are thus obtained directly, and hence the frequency. The mass of S is increased by introducing one or more steel rods into the hollowed body of S, and a new natural frequency is obtained in the same manner as before. The stiffness coefficient may be changed by using different springs. Table I contains typical data. In obtaining the calculated frequency n from $n=(1/2\pi)(k/m)^{\frac{1}{2}}$, where k is the stiffness coefficient of the spring, and m the mass of the oscillating body, one third the mass of the springs (13g) was added to the mass in each case, and k was calculated from the observed frequency with S alone. By direct, though much less precise, determination, k was found to be 2 percent lower than as determined here.

Many other familiar applications have been made, such as to vibrating strings and tuning forks, but the use of the stroboscope in determining the speed of rotation of a Beams top[5] deserves special mention. Speed of rotation as a function of air pressure was obtained, but in

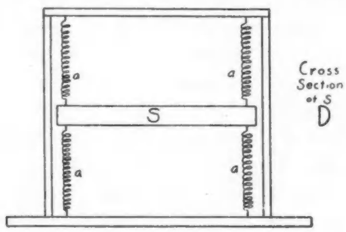

FIG. 3. Oscillating body S, mounted in wooden frame.

this case it was necessary, for the high rates of rotation, to use coincidences for which the top rotated two or more times between successive flashes of the stroboscope. A moderately accurate determination could be made, however.

These typical results indicate the usefulness and attractiveness of the stroboscope for quantitative determination of frequencies of oscillation in experiments and demonstrations. Measurement of the frequencies is very direct, and requires a minimum of time. With the circuit described, the cost is low, the device is easy to operate and the advantages of electrical interruption of light pulses make it reliable and precise.

[5] J. W. Beams, R. S. I. 1, 667 (1930).

To one man science is a sacred goddess to whose service he is happy to devote his life; to another she is a cow who provides him with butter.

—J. VON LIEBIG.

An Undergraduate Experiment in Laue X-Ray Diffraction

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Department of Physics, St. Joseph's College, Philadelphia, Pennsylvania

CRYSTAL analysis by the Laue x-ray diffraction method usually necessitates an x-ray tube of specific design and experimental skill of some maturity. These requirements ordinarily combine to remove such an experiment from the field of undergraduate physics. The experiment here described can readily be incorporated in the undergraduate laboratory. It is qualitative, but simple and direct.

Figure 1 shows the "camera" used.[1] The lead sheets S, C and P are $\frac{1}{8}$-in. thick and 9×7, 6×7.5 and 6×7.5 in. in size, respectively. The screen S contains a 1.5-in. hole through which a brass collimating tube D projects; C is the crystal holder and is separated 2 cm from S; P holds the photographic film and is 5 cm behind C. The collimator D is a $2\frac{3}{8}$-in. length of 1.5-in. brass tubing, on each end of which is soldered a disk with a central hole made with a No. 48 drill. The $\frac{3}{8}$-in. thick brass plate A has a circular hole in which D fits snugly so that the collimator can be worked back and forth, like a telescope draw tube, over a distance of 1.5 cm, thus permitting adjustment of the collimating holes with respect to the crystal. The crystal holder C has a hole, made with a No. 60 drill, in line with the circular apertures of D. The exact location and size of this hole can be determined conveniently by fastening a small square of photographic printing paper on C and exposing it to an incandescent lamp placed in front of the collimator; the well-defined spot on the (developed) print is subsequently drilled through. The crystal is mounted with wax over the hole in C at F. The x-ray film is enclosed in a lightproof envelope which is fastened to P with tape. A housing of $\frac{1}{8}$-in. thick lead, not shown in Fig. 1, is placed over sheets C and P while making exposures, to exclude scattered radiation. The base plate B is of brass of thickness $\frac{1}{8}$ in. In use the camera is mounted about 13 cm from the x-ray tube with the collimator pointed toward the target, the shield S being large enough to cover a rectangular opening cut in the side of the lead box enclosing the tube.

The x-ray tube employed was a commercial Coolidge universal fine-focus type operating off a Snook mechanical rectifier outfit. In use this tube, designed for general medical work, should be "flashed" for a few seconds only since it has no special cooling devices. Fig. 2 shows the central portion of a Laue photograph made with the camera and tube described. The NaCl crystal used, selected with a view to uniformity of thickness, was about 3 mm² and 1.2 mm thick. The net exposure time was 45 min. spread over a total time of 3.5 hr. during which the tube was alternately "flashed" and cooled. The accelerating potential was about 80 kv at approximately 40 ma. Practically any of the various x-ray units now available should prove satisfactory for making photographs like Fig. 2. Photographs were also made with crystals that had been etched, by rubbing on wetted paper, to a thickness of 0.3 mm. This etching process, however, apparently produces strains in the crystal and the resulting Laue patterns exhibit radial streaks (in place of spots) which are usually undesirable.

Some care is necessary in the alignment of the x-ray tube and the 3 circular slits. In the present experiments

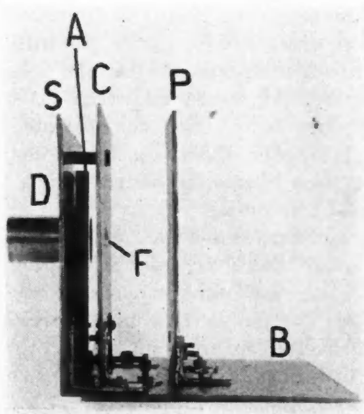

FIG. 1. The Laue "camera."

[1] See also Harnwell and Livingood, ***Experimental Atomic Physics***, p. 335, and Wyckoff, ***The Structure of Crystals***, p. 124, for additional helpful details.

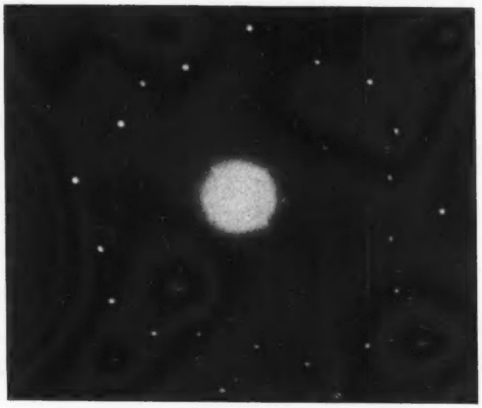

FIG. 2. Laue photograph of NaCl crystal.

the following method was used. The camera, mounted on a tripod base carrying 3 leveling screws, was placed approximately in position to receive the x-ray beam. Plate P was then removed and a focusing type flashlight was clamped in its place and so adjusted that its light, transmitted through the 3 circular diaphragms, produced a well-defined illuminated circular area (about 0.4 cm in diameter) when directed onto the face of the tube target by adjustment of the leveling screws. A further leveling of the camera is then necessary to bring the illuminated circular area exactly on the target focal spot, which is usually visible due to pitting or change in the appearance of the target in the region of electron bombardment. Finally, a fluoroscope was used to check this optical alignment. Care must be exercised in subsequent operations to avoid disturbing the alignment which was found to be somewhat critical. The described method of alignment (1) prevents the danger of over-exposure to x-rays, (2) is safe since the high voltage equipment is not in operation, and (3) is rapid, quite exact, and convenient.

If it is not considered desirable to align the tube and diaphragms, the collimator D can be removed during the exposure. The resulting photograph then will show, in addition to the Laue pattern, a pinhole image of the tube target and, for a qualitative undergraduate experiment, is quite interesting.

Acoustical Interferometers

H. K. SCHILLING
Department of Physics, Union College, Lincoln, Nebraska

IT is the writer's belief that many phenomena of wave motion can be taught more effectively if the first approach to their study is through acoustics rather than optics. In an earlier paper[1] six fundamental experiments on interference were described which can be performed with simple apparatus. To illustrate further the possibilities in the same field, several additional experiments are described here. They demonstrate the essential principles underlying the construction and operation of interferometers with a facility unattainable in optics.

Use is again made of a Galton whistle mounted in a "sound box" as described earlier.[2] The reflectors used are boards, approximately 1×2.5 ft. in size. The partial reflectors, corresponding to half-silvered mirrors in optics, are wire screens in frames of the same dimensions, such as the one used as first reflecting surface of the thin-film model[3] described previously. As sound receivers or detectors, sensitive flames are satisfactory. However, a receiving system consisting of a small-area microphone, an amplifier system, with high pass filter to cut out undesirable laboratory sounds, and an oscilloscope is better.

In the diagrams the partial reflectors are indicated by dashed lines, the total reflectors by heavy lines. All the reflectors are so constructed that they will be upright, perpendicular to the lecture table. This and the fact that they are portable makes it possible to set up and demonstrate a large number of interferometers in a short time. In each diagram the source is represented by a cross, the receiver by a circle. The path of the sound is shown schematically by the lines with arrows. The wavy lines indicate where

[1] H. K. Schilling and Wm. Whitson, Am. Phys. Teacher 4, 27 (1936).
[2] See Fig. 1, Ref. 1.
[3] See Figs. 2, 3, Ref. 1. The reflectivity of these screens can be increased by applying several coats of paint.

blankets or other absorbing screens should be placed to reduce the effects of "stray sounds."

Figure 1 illustrates the acoustic analog of the well-known Michelson interferometer. Since it does not involve refraction and hence does not necessitate the use of a fourth, compensating, plate, both operation and explanation are greatly simplified. The partial reflector should be placed immediately at the aperture of the sound box. Both reflectors a and b may be moved in order to shift "fringes." If they are properly mounted so that during motion they always remain perpendicular to, or at a given angle with respect to, the sound beams, consistent quantitative results can be obtained. The effect of slightly rotating a reflector can also be demonstrated. This interferometer can be set up and properly adjusted in a few minutes—an important desideratum in lecture-table demonstration as well as in the elementary laboratory. For the theory of interferometers the student can be referred to the standard treatises on this subject.[4]

Another advantage of this apparatus is that a given instrument can be readily dismounted and then set up again in different form. This helps the student to understand the fundamental

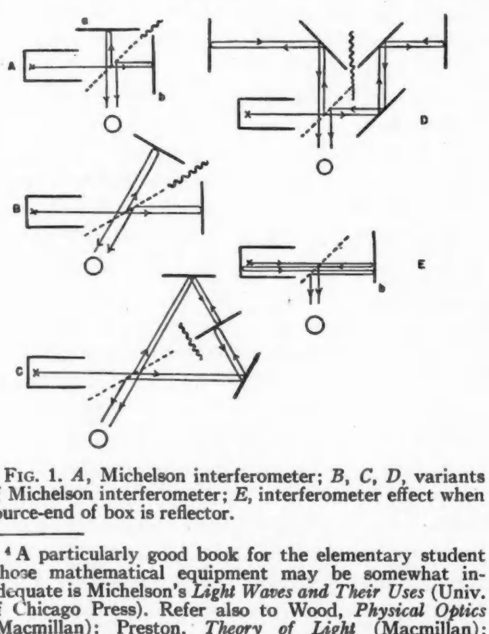

FIG. 1. A, Michelson interferometer; B, C, D, variants of Michelson interferometer; E, interferometer effect when source-end of box is reflector.

[4] A particularly good book for the elementary student whose mathematical equipment may be somewhat inadequate is Michelson's *Light Waves and Their Uses* (Univ. of Chicago Press). Refer also to Wood, *Physical Optics* (Macmillan); Preston, *Theory of Light* (Macmillan); Michelson, *Optical Studies* (Univ. of Chicago Press).

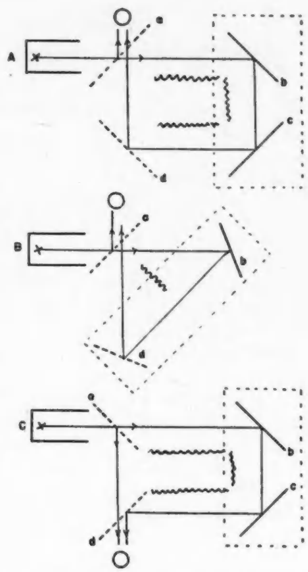

FIG. 2. Interferometers in which one beam travels a much longer path than other.

principles involved since it affords him the opportunity to apply them experimentally to different situations. Thus Figs. 1B, C, D indicate interesting variants of the Michelson interferometer all of which follow the same laws but any one of which might conceivably be more useful than others for a given purpose.

To insure unambiguous results thus far, the source-end of the box must be heavily padded with sound absorbing material. If this is not done and it therefore acts as reflector, interference can be obtained by multiple reflections, without reflector a in 1A, or the corresponding reflectors in B, C and D. This is illustrated by 1E where, even in the absence of a, reflector b can be moved to "shift fringes." Because of multiple reflections between partial reflecting plates this interferometer may be considered to be related to the Fabry-Perot type.

In the Michelson interferometer and the related types just discussed, the acoustical paths of the two interfering beams are nearly equal in length. Those of Fig. 2 are essentially different in that the path of one beam is considerably longer than the other. Partial reflectors are used at d in Fig. 2A and B so that the beams may have more nearly equal intensities at the microphone. The transmittivity of our screen reflectors is much greater than the reflectivity; hence the

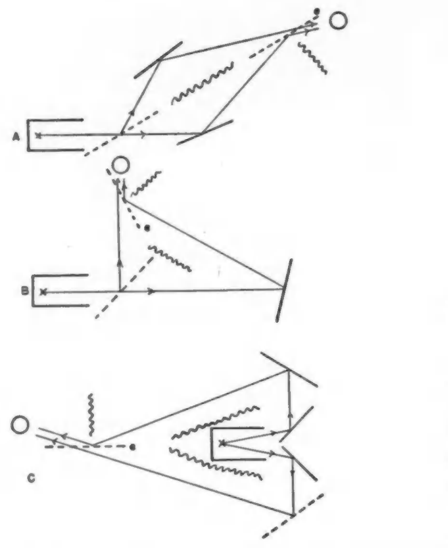

FIG. 3. Interferometers with receiver remote from source.

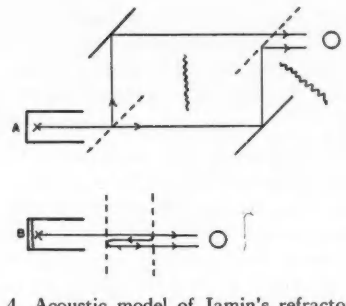

FIG. 4. *A*, Acoustic model of Jamin's refractometer; *B*, Fabry-Perot interferometer.

beam reflected by partial reflector *a* is much less intense than the transmitted beam. This is compensated for by using a partial reflector at, say, *b*. If the length of path *abcd* is very long and the intensity is decreased sufficiently because of this distance factor, a total reflector should be used at all the positions *b*, *c*, *d*. Reflectors *b* and *c* in Fig. 2*A*, and reflectors *b* and *d* in Fig. 2*B*, may advantageously be mounted on a board, indicated by rectangles, to facilitate moving them simultaneously.

Figure 2*C* suggests other possibilities. This particular arrangement differs from Fig. 2*A* in that reflectors *a* and *d* are each turned through 90°, thus necessitating a different position for the receiver. The student will find it instructive to ask himself what would happen if, for instance, *a* and *b* were each turned through 90°, or how the apparatus would have to be changed otherwise if *a* were turned through an angle of 45°, or any other angle.

In Figs. 1, 2*A* and 2*B* the interfering beams are superposed and studied at points near the source, or, at any rate, near the first partial reflector. In Fig. 3 interference takes place at points remote from the source. The diagrams are self-explanatory and need not be discussed further here. Again the student should be challenged to devise other arrangements, or to discuss, for instance, the necessity or desirability of using the partial reflectors *e*, etc.

In Fig. 4*A* we have an acoustical model of Jamin's refractometer; it is, of course, definitely related to Fig. 3*A*.

Figure 4*B* is the equivalent of a Fabry-Perot interferometer. It is important that the end of the box be padded to confine multiple reflections to the space between the two partial reflectors, as indicated in the diagram. The intensity minimums due to destructive interference will not be zero because the transmitted beam is more intense than the reflected one. Likewise the intensity at maximum is not much greater than for minimum. As the distance between partial reflectors is changed, however, there is an unmistakable variation of intensity between maximums and minimums, as predicted by theory.

This subject offers many possibilities for laboratory projects for the more adventurous student. As examples, the writer suggests: (1) to construct and study an echelon; (2) to establish, say, a "standard ten-meter length" by means of sound waves, using Michelson's method; (3) to study "visibility curves" when two sources emitting tones of different pitch are mounted in the box. Other projects will suggest themselves to every teacher.

The writer wishes to acknowledge with thanks the help rendered by his assistant, Mr. Wm. Whitson, in developing these experiments.

DISCUSSION AND CORRESPONDENCE

\ Simple Rule for Directions in Electromagnetic Phenomena

MANY rules have been given for finding directions in the electromagnetic field. Some of the French ones deal with a certain natatorial performance; the usual American ones require the pointing of three fingers, and one must remember not only which hand to use but also which finger is arbitrarily nominated to do what.[1] Instead of these rules the writer has long employed a simple method that even he can remember. Perhaps it would be of interest to others.

The usual conventions as to directions are based upon the behavior of right-hand screws. Now if the thumb of the right hand is pointed outward with the fingers loosely curled, the hand suggests a right-hand screw advancing in the direction pointed out by the thumb and turning round in the direction in which the fingers point round the thumb. (I use this rule when I have to reach around a corner and turn a screw.)

To get the field around a current, therefore, we have only to point the thumb with the current, and the fingers then show us the direction of the magnetic field.

To get the force on a current in a magnetic field, we note further that when the fingers are loosely curled, the first joint or segment of each finger points one way at right angles to the thumb, and the remaining joints point roughly at right angles to both. So we point the thumb with the current, which, we remember, always plays the leading part in the drama, the first joints with the field, which is the second actor, and the last joints then obligingly point out the direction of the resulting force.

For the direction of the electromotive force in a conductor moved across a field, we recall that the reason there is an e.m.f. is that the static electricity in the conductor is carried along with it in its motion and so acquires the properties of a current; the field then pushes sideways on the moving electricity, and this push gives rise to the e.m.f. So we start in the same way as before by pointing the thumb with this current, that is, in the direction of the motion, turn the hand until the first joint of each finger points with the field, and the remaining joints point out the direction of the resulting e.m.f.

Thus a single device, with the help of a little physics, serves in all three cases.

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Ithaca, New York.

[1] Improved rules, the second resembling that offered here, have been given by A. T. Jones, Am. Phys. Teacher 3, 86 (1935), and H. Crew, Am. Phys. Teacher 3, 138 (1935).

Illogic in Textbooks

THE recent article by Witmer and Bushkovitch[1] on the lack of logic in the literature of physics will recall to the minds of readers many additional illustrations which might very profitably be shared with teachers and prospective authors. The first logical demand of Witmer and Bushkovitch, that no more than a necessary minimum number of principles, postulates, or laws be stated, is unanimously or almost unanimously violated in elementary textbooks in the presentation of the so-called motor and dynamo rules. The force on a conductor carrying current in a magnetic field and the electromotive force resulting when magnetic lines of force cut conductors are regularly presented as two distinct natural wonders, which the student is encouraged to keep straight in memory by separate mental feats associated with fists and fingers.

The two phenomena may be summarized by some such single statement as the following. Whenever an electron is impelled by any means to cross a magnetic field it experiences a sideways force urging it to swerve off in a direction at right angles to the field. If the electron chances to be inside a wire, pushed along across the field by an applied potential difference, it is obliged to carry the wire with it in its lateral swerving and we have a motor. In case the containing wire is itself being forcibly carried sideways across the field, the lateral urge upon the electrons carries them *along* the wire, and a generator results.

The fundamental phenomenon of response to a forward push by a sideward detour seems odd, as always, but perhaps the student has seen gyroscopes do something similar. If mnemonic assistance is needed one model will serve instead of two confusingly similar ones. For example, the electrons may be thought of as crossing the earth's magnetic field from east to west, in which case it causes no surprise in our part of the country to find them imbued with something of the buoyancy, the onward-and-upward disposition, ascribed to pioneers.

PAUL KIRKPATRICK

Stanford University,
California.

[1] Am. Phys. Teacher 5, 145 (1937).

A Method of Handling Elementary Laboratory Apparatus

ONE of the difficult problems encountered in the management of the elementary laboratory is that of distributing, collecting, and storing the apparatus. Various solutions have been proposed, which range from allowing the students to help themselves from the supply shelves to installing permanently the equipment on the laboratory tables. There is in this department a system

initiated by the author seven years ago; although its general features have long been known and doubtless used by other departments, the system has proved so successful with us that a brief description of it may be useful.

It is our policy to have the entire laboratory section perform the same experiment simultaneously; hence, except for a few experiments in which the apparatus cost is excessive, twelve sets of apparatus are provided for each pair of students. In view of the quantity of apparatus thus involved, it is desirable to simplify the method of finding and serving equipment so that even a new instructor would have little difficulty in carrying on the work.

To accomplish this end, the apparatus cases are individually lettered and the shelves are numbered. To simplify the search for equipment, a "service book" is provided for each of the major divisions of physics, in which the individual experiments are listed, both by number and by title. For each experiment, detailed instructions are given (Fig. 1) for both serving and storing of the apparatus, and notes are included on the need for stock solutions, transference of multi-use equipment such as calipers and weight sets, and the servicing of special equipment before storing.

The experiment boxes mentioned in Fig. 1, twelve for each experiment, are built to hold all of the small parts of the individual experiment. Construction of the boxes for a given experiment requires, therefore, the assembling of the particular apparatus to determine the dimensions of the boxes. On one end of each box is the experiment number, on the other end, a list of the apparatus which the student is to hand in at the close of the laboratory period (Fig. 2).

Experiment No. 17

Archimedes' Principle

Set up on tables

Balance (Case No. T 1)

Small battery jar (Case No. D 1)

Have available in laboratory the following:

Water

Kerosene

Serve

Boxes for Experiment No. 17 (Case No. B 2)

ADD:

1 vernier caliper (Case No. S 1)

1 set weights (Case No. T 3)

Store

Battery jars in Case No. D 1

Boxes in Case No. B 2

Vernier calipers in Case No. S 1

Leave balances and weights out for Experiment No. 19.

NOTE:

Clean and oil calipers before storing.

FIG. 1.

FIG. 2.

The advantages of this system are obvious: the instructor can distribute the apparatus with a minimum of trouble; the student knows that he is responsible for certain apparatus, and is more apt to check the contents of the box before it is turned in than if the apparatus is loose; the storage space required for a set of boxes is smaller than if the apparatus is piled on shelves.

To install such a system requires a considerable amount of labor but, once started, it requires little attention. A yearly inventory will indicate where replacements are needed.

SANFORD C. GLADDEN

University of Mississippi,
University, Mississippi.

Impact of Elastic Spheres

SOME time ago Professor Harvey B. Lemon[1] called attention to an interesting case of impact in which two elastic spheres having masses in the ratio 3 to 1 collide with velocities that are equal in magnitude but opposite in direction and along their line of centers. In repeating and studying this experiment, which interests everyone who sees it, we found material which seems worth adding to his brief account.

On first thought it seemed just as well to begin by drawing aside the small sphere and allowing it to swing against the larger one at rest. This is simply to begin the cycle (Lemon's Fig. 2) in opposite phase, and then it is not necessary to adjust the balls to the same elevation, as has been done in his Fig. 1, which shows the start. This we did, and then observed that the spheres came back to equal elevations after the first impact, which showed of course that directly after striking together their opposite velocities were numerically equal. After the second impact the large sphere remained perfectly at rest. We tried also the other case suggested, and allowed the large ball to swing against the motionless small one, with the expected result that after two collisions the small ball remained at rest and continued so to speak to sit out the alternate swings of the large one.

In spite of having carried out this experiment, however, I misunderstood it completely. I supposed that the ratio 3 to 1 for the masses was necessary in order that one ball might be still during alternate trips of the other. Of course, if the balls are started with velocities equal in magnitude and opposite in direction, as in Lemon's arrangement, only

the 3 to 1 ratio will serve. But if one starts with either ball in motion, and the other motionless, then what is unique about the 3 to 1 ratio is the opposite velocities of equal magnitude after the first collision. Only the 3 to 1 ratio will give this equality. But with masses in any ratio whatever, the ball originally at rest remains motionless during alternate excursions of the other. Whatever their relative masses, except they be equal, the ball first drawn aside makes two trips to one by the other. This is the most striking property of the experiment as actually seen.

It is easy to predict this result. The equations for the two velocities after impact are

$$\begin{aligned} U_1 &= \{M_1V_1+M_2(2V_2-V_1)\}/(M_1+M_2), \\ U_2 &= \{M_2V_2+M_1(2V_1-V_2)\}/(M_1+M_2), \end{aligned} \quad (1)$$

where V_1 and V_2 are the original velocities, and U_1 and U_2 those after one impact. Let X_1 and X_2 be the velocities after the second impact. We obtain their values by substituting for V_1 the whole expression for U_1, and for V_2, that for U_2, in Eqs. (1); thus, after collecting terms, $X_1=V_1$, $X_2=V_2$.

It seems that the process of colliding is like an operator that, applied to any quantity and then applied to the first result (perhaps better said, applied twice in succession to any quantity), restores the quantity to its original form or value. Such is the operation of applying the minus sign or, more aptly, of taking the reciprocal. The reciprocal of the reciprocal is the original, and after two collisions we have the original situation. An apparent exception is given by balls of equal mass, for with them we have the original velocities 0 and V_1 after one collision. But unity is unique in its reciprocal also, since the reciprocal of 1 is 1, and we do not have to repeat the process in order to recover the original number. However, after one collision with equal balls the velocities are interchanged. It seems that the perfect analogy is to take the negative reciprocal, for here unity is not exceptional and with any number two operations are necessary and sufficient to bring back the original. If the spheres are A and B, two collisions are necessary before A has its original speed, even if A and B are of equal mass. We are reminded here of a fundamental distinction in statistics. Now however, we cannot evade a further question. Two collisions being so much like two operations of taking the negative reciprocal, is one collision like one such operation? But perhaps " 'twere to consider too curiously to consider so."

The physical basis of the process is apparent after a little reflection. At the very start, that is when one ball is first released, a certain amount of energy is represented in the system, which is conserved. At every impact momentum is constant also, and these two facts give us two equations, one involving the squares of velocities, and these equations give two values for each new velocity, U_1 and U_2, after the impact. We reject one pair of values because they represent the original velocities. They would imply that one sphere passed through the other unhindered. To get X_1 and X_2, the velocities after the second impact, we use the same energy and the same momentum as before and hence we get the same values for the velocities after the second collision as after the first. But we reject now the two which we first preserved, and preserve those which we had before the first collision. We reject and preserve alternately. It follows then that after two collisions along the line of centers the original situation reappears, whether one ball only, or both of them, were first set in motion.

In performing this experiment it is well to see that the spheres when at rest are in contact, with the supporting wires of each sphere in a vertical plane. Also, the motion should be along the line of centers and the periods of the two should be the same. The amplitudes of the swings should not be too large, for otherwise the periods will be unequal and successive collisions will not occur at the same place.

W. W. Sleator

University of Michigan,
Ann Arbor, Michigan.

[1] Am. Phys. Teacher 3, 36 (1935).

Activities of the Kentucky Chapter

CONVENING simultaneously with the annual State Educational Conference and the State Association of Colleges and Secondary Schools, on October 30, the Kentucky Chapter of the American Association of Physics Teachers presented the following program in the physics lecture room of the University of Kentucky:

Apparatus for Demonstrations. T. M. Hahn, *University of Kentucky.*
Problem Sections. D. M. Bennett, *University of Louisville.*
The Physics of General Science. J. S. Cantwell, *Halleck Hall.*
Meteorology as a Project. H. C. Mitchell, *First Creek High School.*
The Future of Physics in the High School. W. K. Evans, *Bryan Station High School.*

Each year the Department of Extension of the University of Kentucky gives competitive examinations to secondary school students with awards granted as follows:

1. To the school whose student makes the highest score, a bronze plaque, awarded by the Kentucky Chapter of the American Association of Physics Teachers.

2. To the student having the highest score, a medal, awarded by Lambda Chapter of Sigma Pi Sigma.

3. Other awards to successful competing schools in the form of physics apparatus.

TEACHING AIDS

Exhibits

Celestialite Glass. *Gleason-Tiebout Glass Co.* (200 Fifth Ave., New York), gratis. Sample fragments of Celestialite lighting glass, showing its construction.

Metric Ruler. *Metric Association* (Pottsville, Pa.), 50 cts. per 100. A 15-cm wooden pocket ruler.

Charts and Posters

Stromberg Carburetors. Chart No. 10–9. 81×105 cm. *Bendix Products Corp.* (401 Bendix Dr., South Bend, Ind.), gratis. Cross-sectional diagrams of 4 types of carburetors, in colors and with explanatory legends.

Carrier Psychrometric Chart. 41×46 cm. *Carrier Corp* (850 Frelinghuysen Ave., Newark, N. J.), gratis. A psychrometric chart, with instructions for use and worked examples. Available in quantities for distribution to engineering classes.

Irradiated Evaporated Milk. 60×90 cm. *Irradiated Evaporated Milk Institute* (307 N. Michigan Ave., Chicago), gratis. Shows the steps in the process of preparing evaporated milk irradiated with ultraviolet light.

Weights and Measures Charts of the National Bureau of Standards. Superintendent of Documents, Government Printing Office (Washington):

The international metric system, Misc. Pub. BS. M3 (1936). 71×110 cm, 30 cts. Shows graphic comparisons, in colors, between the various metric and English units.

Standard time conversion chart, Misc. Pub. BS, M84 (1931). 20×26 cm, 10 cts. A heavy cardboard chart with a movable disk, from which the time at any other part of the earth may be read directly.

Time zone map of the United States, Misc. Pub. NBS, M155 (1936). 50×75 cm, 10 cts. Shows location of cities and time-zone boundaries, in colors.

Motion Picture Films

Approved by the Underwriters. 16 mm sound film, 40 min. *Underwriters Laboratories* (Chicago), loaned gratis. Testing of devices to prevent loss of life and property from fire, accident, and theft.

Motion Picture Films of the Bureau of Mines. 26 p., 9×19 cm. *U. S. Bureau of Mines Exp. Sta.* (4800 Forbes St., Pittsburgh, Penna.), gratis. The 1937 catalog of silent films which may be borrowed from the various distributing centers of the Bureau of Mines. Some of the subjects listed are: Story of a spark plug (2 reels, recently revised); Story of lubricating oil (2 reels); Carbon monoxide, the unseen danger; (1 reel); Valves, their manufacture and uses (3 reels); Wildwood, a hundred percent mechanized mine (3 reels); Automobile lubrication (1 reel); Metals of a motor car (2 reels); Nickel (2 reels); Silver (3 reels); Construction and operation of internal combustion engines (2 reels, recently revised); Story of a storage battery (2 reels); Story of the gasoline motor (3 reels).

Pamphlets and Trade Bulletins

Publications on the Telephone. The following pamphlets may be obtained gratis from the American Telephone and Telegraph Co., Information Dept., 195 Broadway, New York, or from any local Bell company:

The Magic of Communication, by John Mills. 38 p., 40 fig., 15×22 cm. An excellent, simple description of the operation of telephone and telegraph systems.

The Birth and Babyhood of the Telephone. 47 p., 29 fig., 13×19 cm. An interesting account of the early history of the telephone, written by Alexander Graham Bell's assistant, Thomas A. Watson.

The Telephone in America. 67 p., 104 fig., 15×23 cm. A well illustrated account of the history, organization, aims, research, and numerous other activities of the Bell System.

Fuseology. 18 p., 28 fig., 18×23 cm. *Bussman Mfg. Co., Sales Promotion Dept.* (University at Jefferson, St. Louis), gratis. Comprehensive and valuable information on fuses and on how to avoid fusing troubles.

Stromberg Instruction Bulletins. 130 p., 21×27 cm. *Bendix Products Corp.* (401 Bendix Dr., South Bend, Ind.), gratis. Loose-leaf bulletins describing the construction and servicing of various types of Stromberg carburetors. Illustrated with cross-sectional diagrams.

Zenith Carburetors. Form 2016. 12 p., 20 fig., 21×27 cm. *Zenith Carburetor Co.* (696 Hart Ave., Detroit, Mich.), gratis. An excellent, simple description of the construction and operation of Zenith carburetors. Good diagrams.

Celestialite. 19 p., 19 fig., 18×25 cm. *Gleason-Tiebout Glass Co.* (200 Fifth Ave., New York), gratis. A brief non-technical description of the uses of this lighting glass.

Catalog and Blueprints of Heating Specialties. *The Trave Co.* (La Crosse, Wis.), gratis. Descriptions and diagrams of various types of radiator valves and steam traps. Cut samples of valves and traps can be obtained from this company at cost.

The Story of Evaporated Milk. 32 p., 17 fig., 15×23 cm. *Evaporated Milk Association* (307 N. Michigan Ave., Chicago), gratis. This pamphlet, prepared for classroom use, includes a good illustrated description of the process of transforming raw milk into evaporated milk.

Publications on Irradiated Evaporated Milk. *Irradiated Evaporated Milk Institute* (307 N. Michigan Ave., Chicago), gratis. Various publications, including one on "The Story of Irradiated Evaporated Milk," which describe the appli-

cation of ultraviolet radiation to increase the vitamin D content of foods.

Free Publications of the Bureau of Mines. *U. S. Bureau of Mines* (Information Div., Washington, D. C.). The following free publications are among those of possible interest to physicists:

About helium, IC6745. 46 p. Tells of the discovery, properties, sources, production, and uses of helium, and of the work of the Bureau's cryogenic laboratory. Bibliography.

Note on the use of ultraviolet lamps in mines for rapid determination of scheelite in ores by fluorescence. IC6873, 3 p., 1 fig., A simple description of the application of fluorescence in tungsten-mining industry.

Absorption of nitrogen by steel. RI3076. 8 p., 2 fig.

Note on copper-constantan thermocouple calibration below 0° C. RI3077. 7 p., 1 fig.

Wanted: more detailed reports on electrical accidents. IC6046. 9 p. Indicates the information needed for adequate reports.

Mine safety as affected by electrification. IC6052. 3 p.

Radium. IC6312. 55 p., 1 fig.

A comment upon present-day geophysics. IC6496. 5 p. Simple explanation of the basic aims and methods of geophysics.

Tourmaline. IC6539. 8 p.

Weights and Measures Publications of the Bureau of Standards. *National Bureau of Standards* (Washington). The following publications may be obtained gratis:

Standards of length, mass, and time. LC449 (1935). General information.

Metric and English distance equivalents for athletic events. LC376 (1933). Gives the yard distances of various track events, with metric equivalents.

Weights and measure publications of general interest. IC492 (1937). Describes 23 publications that can be obtained at small cost from the Superintendent of Documents. Some of the titles are: "Standard density and volumetric tables" (15 cts.); "Copper wire tables" (20 cts.); "Sundials" (5 cts.); "History of the standard weights and measures of the U. S. (15 cts.); "Units of weight and measure-definition and tables of equivalents" (15 cts.).

DIGEST OF PERIODICAL LITERATURE

General Physics

Newton's Third Law of Motion. V. F. Lenzen; *Isis* **27**, 258, Aug., 1937. An examination of Newton's discussion of the third law of motion in the *Principia* (Cajori's revision, p. 13) reveals an ambiguity which appears to have been overlooked by commentators. The third law states,

> "To every action there is always opposed an equal reaction; or, the mutual actions of two bodies upon each other are always equal, and directed to contrary parts."

In illustration,

> "Whatever draws or presses another is as much drawn or pressed by that other. If you press a stone with your finger, the finger is also pressed by the stone."

Newton's statement of the law and his illustrations suggest that action and reaction act on different bodies. If the finger presses a stone, the stone exerts a force equal in magnitude and opposite in direction, the reaction, on the finger. This generally accepted interpretation of the third law is confirmed in the first part of the scholium which concludes the account of the laws of motion. In the first part, Newton describes experiments on the impact of bodies in which action and reaction act on different bodies. He also presents arguments to prove that the third law applies to attractions.

Continuing the discussion of the third law he says,

> "And as those bodies are equipollent in the impact and reflection, whose velocities are inversely as their innate forces, so in the use of mechanic instruments those agents are equipollent, and mutually sustain each the contrary pressure of the other, whose velocities, estimated according to the determination of the forces, are inversely as the forces. So those weights are of equal force to move the arms of a balance, which during the play of the balance are inversely as their velocities upwards and downwards; that is, if the ascent or descent is direct, those weights are of equal force, which are inversely as the distances of the points at which they are suspended from the axis of the balance."

This quotation expresses a second concept of action and reaction. The two bodies, which are suspended from the arms of the balance, exert forces on the same body and therefore are not action and reaction as previously defined. If body A exerts a force on the balance, the reaction to it is the force that the balance exerts on A; similarly for the body at the other end of the balance. The forces acting on the balance may be called *applied force* and *resistance.* This language is employed by Newton,

> "The power and use of machines consist only in this, that by diminishing the velocity we may augment the force, and the contrary; from whence, in all sorts of proper machines, we have the solution of this problem: To move a weight with a given power, or with a given force to overcome any given resistance."

After discussing such examples he concludes,

> "I was aiming only to show by those examples the great extent and certainty of the third law of motion."

In order to make Newton's account of the third law self-consistent it is necessary to assume that he used action and reaction in two senses: as *force* and *counter-force* which act on different bodies, and as *applied force* and *resistance* which act on the same body.

The twofold meaning of action and reaction is exhibited in some commentaries. Thus L. Bloch [*La Philosophie de Newton*] confuses counter-force and resistance. In one passage he uses reaction in the sense of resistance which acts on the same body as the action. Thomson and Tait [*Treatise on Natural Philosophy* (1879), vol. 1, part 1, p. 247] introduce the third law with a statement and examples which imply that action and reaction act on

different bodies. But in their subsequent discussion they employ the principle that the counter-activity of the resistance is equal and opposite to the activity of the agent. Since the resistance acts on the same body as the agent, these authors contribute to the ambiguity in the concept of reaction.

Contemporary writers on mechanics usually state explicitly that action and reaction act on different bodies. Some of the writers who accept this restricted meaning of action and reaction also explain equilibrium in terms of the third law. They thereby confuse reaction and resistance. A balance is in equilibrium if the moment of the applied force is equal and opposite to the moment of the resistance. Both moments act on the balance and it is incorrect to derive their equality from a third law which assumes that action and reaction act on different bodies. The writer thinks that the third law should be restricted to forces that act on different bodies, and that equilibrium should be treated as a special case of the second law.

History and Biography

Frederic E. Ives. H. C. Richards; *Science* **86**, 340–1 (1937). Frederic Eugene Ives died on May 27, 1937, at his home in Philadelphia. His achievements were many and important, and were all the more remarkable in that he was essentially a self-educated man; he never attended a formal school after he was twelve years of age. He was born on February 17, 1856, on a farm near Litchfield, Connecticut. When he was twelve his father died, which interrupted his schooling and forced him to earn his living. After a brief and unsatisfactory experience as clerk in a country store he became apprentice in the printing office of the *Litchfield Enquirer*. Here he became interested in picture making both by the process of wood engraving and the art of photography. The limitations of the former method impressed him with the advantages of the photographic process and led him later to attack the problem of photoengraving. His spare time, which was little, was spent in experiments in photography, his first camera being made, as he tells us, of a cigar box and a spectacle lens.

In 1847 he was employed in the photographic laboratory of Cornell University, where he remained for four years. His contact with this institution doubtless stimulated his interest in the scientific aspect of his work and inspired his inventive genius. It is to these years that we refer his first notable achievement, the invention of the first commercially successful halftone process of photoengraving. The success of this invention, which has revolutionized the art of illustration, led him to connect himself with a printing establishment in Philadelphia in which he worked as photoengraver, maintaining at his home a private laboratory where he busied himself incessantly in numerous inventions. Here in 1886 he developed the cross-line screen method of halftone reproduction which is now universally used. Here also he carried out extensive experiments in color photography and color reproduction by the trichromatic process; his brilliant success in the solution of this problem by the invention in 1892 of the photochromoscope won for him recognition in the scientific world. This was perhaps his greatest achievement, and one in which he continued to work throughout his life, developing improved methods and new applications. But his work also bore fruit in many other ways. Numerous devices, mostly in the field of optics, are due to his ingenuity. Over seventy patents were taken out for his inventions, and as many more could easily have been obtained.

Ives was a member of many scientific organizations. The chief characteristic of his method of work is that it was firmly based on true scientific principles. His cross-line halftone process was worked out with a thorough understanding of the optical principles involved in lens aperture, line spacing, etc. His work in color reproduction shows a complete grasp of the trichromatic theory of Young, Helmholtz and Maxwell which was completely lacking with most of the other experimenters in this field; consequently, all subsequent work in color printing and color photography is based upon fundamental investigations.

Unfortunately he did not reap the proper material reward for his ingenuity. Some of his inventions—notably his halftone process—were unprotected by patents. Others were the subject of costly litigation or were infringed upon by his competitors. Fortunately for himself and for the world he was a type of man who—to use his own words—"will pursue his course through any amount of poverty and hardship and indifference, thinking much more about his work than about any material reward which it may bring."

Appointment Service

REPRESENTATIVES of departments or of institutions having vacancies are urged to write to the Editor, Columbia University, for additional information concerning the physicists whose announcements appear here or in previous issues. *The existence of a vacancy will not be divulged to anyone without the permission of the institution concerned.*

20. Ph.D. Univ. of Minnesota; S.B., S.M., M. I. T.; 1 yr. grad. work, Univ. of Iowa. Age 38, married, 2 children, 17 yr. teaching experience in universities, colleges and technical schools, including 10 yr. head of department. Interested in progressive undergraduate and graduate teaching and research, including mathematical physics.

21. M.S. Kansas State. Age 38, married, 2 children. Research, acoustics. Experienced in laboratory maintenance, design and construction of apparatus, and writing of manuals. Desire teaching or industrial laboratory employment.

Author Index to Volume 5

In this index are listed the names of the authors and the titles of their articles. Abstracts of papers read at the annual meeting and articles that appear in the "Digest of Periodical Literature" section of the journal are not listed; they are indexed in the Analytic Subject Index.

Analytic Subject Index to Volume 5

The titles of articles are disregarded, the entries being based on analyses of the contents of the original articles. Entries marked (D) refer to digests appearing under "Digests of Periodical Literature" and to abstracts of papers read at A.A.P.T. meetings; entries marked (R) refer to reviews appearing under "Recent Publications" and "Teaching Aids."